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MATHEMATICS

ON A REPRESENTATION OF ADDITIVE OPERATOR SET FUNCTIONS

By M. A. NEUMARK

(Communicated by A. N. Kolmogoroff, Member of the Academy, 15.VIII.1943)

Let X be an arbitrary set; $\mathfrak{S}$, a system of subsets Δ of X containing X and the void set O and with every two sets containing their sum and difference (hence, also their intersection). Let further $\mathfrak{H}$ be a unitary space of arbitrary dimension, and $B(\Delta)$ an additive set function on $\mathfrak{S}$ satisfying the following conditions:

1°. For any $\Delta \in \mathfrak{S}$ $B(\Delta)$ is a positive Hermitian operator in $\mathfrak{H}$.

2°. $B(0)=0$, $B(X)=1$, where 0 and 1 in the right are the null-operator and the unit-operator, respectively.

For the case, when X is the set of all real numbers, $\mathfrak{S}$, the system of all Borel sets on X, we have proved [(1), Theorem 1], that every such function $B(\Delta)$ can be represented in the form $B(\Delta)x = PE(\Delta)x$, $x \in \mathfrak{H}$, where $E(\Delta)$ is an orthogonal spectral function (i. e. a projection, additively depending on Δ and satisfying 2°) in a unitary space $\mathfrak{H}' \supset \mathfrak{H}$ and P is a projection in $\mathfrak{H}'$ onto $\mathfrak{H}$.

A considerably more simple proof of this result and besides for a more general case, when X is the character group of a locally-compact commutative group, was further given in (2) (Corollary 1). In this proof we have used some results concerning positive definite operator functions on a group.

In the present note we give a still more simple proof of this theorem and besides in its most general form, when X is an arbitrary set, and $\mathfrak{S}$ an arbitrary system satisfying the conditions enumerated above.

Theorem. *Every additive operator function $B(\Delta)$ in the unitary space $\mathfrak{H}$ can be represented in the form $B(\Delta)x = PE(\Delta)x$, $x \in \mathfrak{H}$, where $E(\Delta)$ is an orthogonal spectral function in a unitary space $\mathfrak{H}' \supset \mathfrak{H}$ and P is a projection in $\mathfrak{H}'$ onto $\mathfrak{H}$.*

Proof. Consider the set $\mathfrak{M}$ of finite formal sums $z = \sum_k x_k \Delta_k$, where $x_k \in \mathfrak{H}_k$, $\Delta_k \in \mathfrak{S}$, and $\Delta_k \Delta_j = 0$ for $k \neq j$. Two such sums $z = \sum_k x_k \Delta_k$ and $z' = \sum_k x'_k \Delta'_k$ we consider as equal, if one of them can be obtained from the other by the following operations: (a) replacement of a group $x\Delta_{k_1} + \ldots + x\Delta_{k_p}$ of terms by one term $x(\Delta_{k_1} + \ldots + \Delta_{k_p})$, (b) the inverse operation of replacement of $x(\Delta_{k_1} + \ldots + \Delta_{k_p})$ by $x\Delta_{k_1} + \ldots + x\Delta_{k_p}$; (c) cancellation or addition of terms $O \cdot \Delta$.

Using the operations (a), (b) and (c) we can put every two elements $z, w \in \mathfrak{M}$ in the form $z = \sum_k x_k \Delta_k$, $w = \sum_k y_k \Delta_k$, where the sets Δ_k and their number are the same. Put then

$$\lambda z = \sum_k \lambda x_k \Delta_k, \quad z + w = \sum_k (x_k + y_k)\Delta_k, \quad (z, w) = \sum_k (B(\Delta_k)x_k, y_k), \qquad (1)$$

where λ is a scalar. This definition does not depend on the choice of the representation of z and w with the same Δ_k. In fact, the operations (a), (b), (c) simultaneously applied to the Δ_k in z and w lead to the same operations in λz and $z + w$.

As to (z, w), the assertion follows from the additiveness of $B(\Delta)$ and the condition $B(0) = 0$. Put $\|z\| = \sqrt{(z, z)}$ and call z and w equivalent if $\|z - w\| = 0$. Then $\mathfrak{M}$ will be decomposed into equivalence classes ξ. Evidently, to the first two operations (1) there will correspond operations with classes; moreover, we can introduce scalar multiplication of classes, if we put $(\xi, \eta) = (z, w)$ for $z \in \xi$, $w \in \eta$.

Then $(\xi, \xi) = 0$ implies $\xi = 0$, so that the set Ξ of all classes ξ becomes a (incomplete in general) unitary space. Completing it with respect to (ξ, η), we get a unitary space $\mathfrak{H}'$.

The elements x of the initial space $\mathfrak{H}$ we can identify with the elements of $\mathfrak{M}$ of the form $z = xX$; if $z = xX$, $w = yX$, then by 2° $(z, w) = (B(X)x, y) = (x, y)$. Identifying further $x \in \mathfrak{H}$ with the equivalence class containing xX, we can therefore consider $\mathfrak{H}$ as a subspace of $\mathfrak{H}'$.

Let now $\Delta \in \mathfrak{S}$; introduce an operator $F(\Delta)$ in $\mathfrak{M}$ by $F(\Delta) \sum_k x_k \Delta_k = \sum_k x_k \Delta\Delta_k$.

Evidently $F(\Delta)z$ does not depend on the representation of z, is linear and $(F(\Delta))^2 = F(\Delta)$. Moreover,

$$(F(\Delta)z, w) = \sum_k (B(\Delta\Delta_k)x_k, y_k) = (z, F(\Delta)w) \text{ and } \|F(\Delta)z\| \leqslant \|z\|.$$

These relations show that $F(\Delta)$ can be considered as an operator $E(\Delta)$ in $\mathfrak{H}'$; they imply also that $E(\Delta)$ is a projection in $\mathfrak{H}'$. Its additiveness with respect to Δ follows from the evident additiveness of $F(\Delta)$. Moreover, it is evident that $E(X) = 1$, $E(0) = 0$, so that $E(\Delta)$ is an orthogonal spectral function.

For elements $z = xX$, $w = yX$ we have: $(F(\Delta)z, w) = (B(\Delta)x, y)$, or, passing to the corresponding equivalence classes and identifying them, as before, with x and y:

$$(E(\Delta)x, y) = (B(\Delta)x, y) \text{ for } x, y \in \mathfrak{H}. \qquad (2)$$

Let P be the projection in $\mathfrak{H}'$ onto $\mathfrak{H}$; then $P\eta \in \mathfrak{H}$ for any $\eta \in \mathfrak{H}'$, hence by (2) $(E(\Delta)x, P\eta) = (B(\Delta)x, P\eta)$, $(PE(\Delta)x, \eta) = (PB(\Delta)x, \eta) = (B(\Delta)x, \eta)$, because $B(\Delta)x \in \mathfrak{H}$. Hence $PE(\Delta)x = B(\Delta)x$, completing the proof of the theorem.

Remark. The space $\mathfrak{H}'$ constructed in the proof, is minimal in the sense that the orthogonal complement $\mathfrak{H}' - \mathfrak{H}$ contains no subspace, invariant with respect to all $E(\Delta)$ and different from (0). In fact, let $\mathfrak{H}_1$ be any minimal subspace and $^*E_1(\Delta)$ the corresponding orthogonal spectral function in it satisfying the relation $P_1 E_1(\Delta)x = B(\Delta)x$; $x \in \mathfrak{H}$, where P_1 is the projection in $\mathfrak{H}_1$ onto $\mathfrak{H}$.

Then $\mathfrak{H}_1'$ is the closure of the set of all finite sums $x = \sum_k E_1(\Delta_k) x_k$, where $x_k \in \mathfrak{H}$, $\Delta_k \in \mathfrak{S}$; $\Delta_k \Delta_j = 0$ for $k \neq j$. To any such sum we put in correspondence the sum $z = \sum_k x_k \Delta_k$. Then for $y = \sum_k E_1(\Delta_k) y_k$, $w = \sum_k y_k \Delta_k$ we get:

$$(x, y) = \sum_k (E_1(\Delta_k) x_k;\ E_1(\Delta_k) y_k) = \sum_k (E_1(\Delta_k) x_k,\ y_k) =$$

$$= \sum_k (E_1(\Delta_k) x_k, P_1 y_k) = \sum_k (P_1 E_1(\Delta_k) x_k,\ y_k) =$$

$$= \sum_k (B(\Delta_k) x_k,\ y_k) = (z, w).$$

If we therefore identify every x with the corresponding z, then $\mathfrak{H}_1'$ will coincide with $\mathfrak{H}'$. Moreover, in virtue of

$$E_1(\Delta) x = \sum_k E_1(\Delta) E_1(\Delta_k) x_k = \sum_k E_1(\Delta \Delta_k) x_k$$

the operator $E_1(\Delta)$ will then coincide with $E(\Delta)$.

Hence, *for a given $B(\Delta)$ the space $\mathfrak{H}'$ and the orthogonal spectral function $E(\Delta)$ constructed in the proof of the theorem can be considered as a model of all minimal spaces $\mathfrak{H}_1'$ and orthogonal spectral functions in them, satisfying the condition* $P_1 E_1(\Delta) x = B(\Delta) x$, $x \in \mathfrak{H}$.

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MATHEMATICS

SEVERAL SERIES CONTAINING PRIMES AND ROOTS OF $\zeta(s)$

By A. I. POPOV

(Communicated by I. M. Vinogradow, Member of the Academy, 11. VIII. 1943)

While studying the problem of the distribution of primes I have by the way obtained a great number of rather interesting series connected with the roots of Riemann's function $\zeta(s)$ (or other series of Dirichlet). The most typical formula of this kind is

$$\sum_{n>x} \Lambda(n) \frac{\left\{\frac{n}{x}\right\} - \left\{\frac{n}{x}\right\}^2}{n^2} = \frac{2-\lg 2\pi}{x} + \sum_{\rho} \frac{x^{\rho-2}}{\rho(\rho-1)} + \\ + \sum_{k=1}^{\infty} \frac{k+1-2k\zeta(2k+1)}{2k(k+1)(2k+1)x^{2k+2}}. \tag{1}$$

Here, as usual, $\Lambda(n)$ is a well-known function of Mertens; $\{z\}$ denotes the fractional part of z; $\sum_{\rho}$ is extended over all complex roots ρ of the function $\zeta(s)$; $x \geqslant 1$; the series on the left-hand side extends over all integer n larger than x and presents a continuous function of x. The following approximate relation may also be inferred from (1):

$$\sum_{n>x} (\Lambda(n)-1) \frac{\left\{\frac{n}{x}\right\} - \left\{\frac{n}{x}\right\}^2}{n^2} = \sum_{\rho} \frac{x^{\rho-2}}{\rho(\rho-1)} + O\left(\frac{1}{x^2}\right). \tag{2}$$

Let us now, according to custom, denote the integral part of x by $[x]$ and put $\zeta(s)(1-2^{1-s}) = \zeta_1(s)$.

We may then (for $x > 1$) write:

$$\sum_{n>x} \frac{\Lambda(n)}{n^2} \left(\frac{1-(-1)^{\left[\frac{n}{x}\right]}}{2} + (-1)^{\left[\frac{n}{x}\right]} \left\{\frac{n}{x}\right\} \right) = \\ = \frac{1-\lg\frac{\pi}{2}}{x} - \sum_{\rho} \frac{x^{-1-\rho}}{\rho} + \sum_{k=1}^{\infty} \frac{\zeta_1(2k+1)-k-1}{(2k+1)(k+1)x^{2k+2}}. \tag{3}$$

Here the roots ρ of the Riemann function are arranged in the increasing order of their moduli, and the convergence of $\sum_{\rho}$ is apparently non-uniform.

Accordingly, the series on the left-hand side of (3) has its continuity broken at points $x = p^h$, where p is a prime number, $h = 1, 2, 3, \ldots$ Adopting Riemann's hypothesis concerning the roots of $\zeta(s)$, that is putting $\rho = \frac{1}{2} \pm \gamma_n i$, we shall instead of (1) obtain:

$$\sum_{n>x} \Lambda(n) \frac{\left\{\frac{n}{x}\right\} - \left\{\frac{n}{x}\right\}^2}{n^2} = \frac{2 - \lg 2\pi}{x} - \frac{2}{x^{3/2}} \sum_{n=1}^{\infty} \frac{\cos(\gamma_n \log x)}{\frac{1}{4} - \gamma_n^2} + \sum_{k=1}^{\infty} \frac{k+1-2k\zeta(2k+1)}{2k(k+1)(2k+1)x^{2k+2}}. \tag{4}$$

The proof of such relations as represented by (1) is very easily carried out on the basis of the theory of Dirichlet series.

The relations of this kind may serve, besides, to determine the numerical value of a great many series whose terms are functions of primes.

We find, for instance, that

$$\sum_{p=3,5,7,11,\ \ldots} \frac{(-1)^{\frac{p-1}{2}} \lg p}{p - (-1)^{\frac{p-1}{2}}} = \lg \frac{8}{\pi} - 2C - \sum_{k=1}^{\infty} \frac{\zeta(2k+1)}{(2k+1)2^{4k}}. \tag{5}$$

Here C is Euler's constant, and the sum on the left side extends over all odd primes.

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MATHEMATICS

EQUATIONS, LINEAR IN JACOBIANS, ON WHICH ARE DISTRIBUTED GENERALIZED-JACOBI, THE JACOBI SYSTEMS OF LINEAR EQUATIONS IN MANY FUNCTIONS, AND THE SYMBOLIC FORMS OF THE CANONICAL TYPE CONNECTED WITH THEM

By G. V. PFEIFFER, Member of the Academy of Sciences of the Ukrainian SSR

Let us consider the generalized-Jacobi system of the q-class:

$$\left.\begin{aligned} Xp_{1j} &= Z_1^j - \sum_{i=1}^{q} X_i^j p_{1i} \\ Xp_{2j} &= Z_2^j - \sum_{i=1}^{q} X_i^j p_{2i} \\ &\cdots\cdots\cdots\cdots \\ Xp_{kj} &= Z_k^j - \sum_{i=1}^{q} X_i^j p_{ki} \\ j &= q+1,\ q+2,\ \ldots,\ n \end{aligned}\right\} \tag{1}$$

which is the same as

$$\left.\begin{aligned} p_{\sigma j} &= Z_\sigma^j - \sum_{i=1}^{q} X_i^j p_{\sigma i},\quad j = q+1,\ q+2,\ \ldots,\ n \\ \sigma &= 1,\ 2,\ \ldots,\ k, \qquad X \equiv 1 \end{aligned}\right\} \tag{2}$$

and the equation linear in Jacobians

$$\left.\begin{aligned} J_n &\equiv \sum_{\nu} \sum_{\sigma_1 \ldots \sigma_\nu}^{k} \sum_{\tau_1 \ldots \tau_\nu}^{n} M_{\tau_1 \ldots \tau_\nu}^{\sigma_1 \ldots \sigma_\nu} \frac{\partial (z_{\sigma_1} \ldots z_{\sigma_\nu})}{\partial (x_{\tau_1} \ldots x_{\tau_\nu})} = 0 \\ \nu &= 0,\ 1,\ \ldots,\ k, \qquad n > k \\ \nu &= 0,\ 1,\ \ldots,\ n, \qquad k > n \\ \nu &= 0,\ 1,\ \ldots,\ k = n, \quad k = n \\ & \qquad\qquad M \neq 0 \end{aligned}\right\} \tag{3}$$

For $q = n-1$ the system (1)—(2) becomes the Jacobi system.
In accordance with the values of q:

$$q = n-1,\quad n-2,\ \ldots,\ n-\rho \tag{4}$$

in the system (1)—(2), the equation (3) will be written down as follows:

$$J_n \equiv J_{n-1} + G_n = 0, \tag{5}$$

$$J_n \equiv J_{n-2} + G_{n-1} + G_n = 0 \tag{6}$$

$$\cdots\cdots\cdots\cdots$$

$$J_n \equiv J_{n-\rho} + G_{n-\rho+1} + \ldots + G_{n-1} + G_n = 0, \tag{7}$$

where

$$G_n \equiv \sum_{\nu} \sum_{\sigma_1 \ldots \sigma_\nu \sigma_{\nu+1}}^{k} \sum_{\tau_1 \ldots \tau_\nu}^{n-1} M^{\sigma_1 \ldots \sigma_\nu \sigma_{\nu+1}}_{\tau_1 \ldots \tau_\nu n} \frac{\partial (z_{\sigma_1}, \ldots, z_{\sigma_\nu}, z_{\sigma_{\nu+1}})}{\partial (x_{\tau_1}, \ldots, x_{\tau_\nu}, x_n)}, \tag{8}$$

$$G_{n-1} \equiv \sum_{\nu} \sum_{\sigma_1 \ldots \sigma_\nu \sigma_{\nu+1}}^{k} \sum_{\tau_1 \ldots \tau_\nu}^{n-2} M^{\sigma_1 \ldots \sigma_\nu \sigma_{\nu+1}}_{\tau_1 \ldots \tau_\nu n-1} \frac{\partial (z_{\sigma_1}, \ldots, z_{\sigma_\nu}, z_{\sigma_{\nu+1}})}{\partial (x_{\tau_1}, \ldots, x_{\tau_\nu}, x_{n-1})}, \tag{9}$$

$$\cdots\cdots\cdots\cdots\cdots\cdots\cdots\cdots\cdots\cdots\cdots\cdots$$

$$G_{n-\rho+1} \equiv \sum_{\nu} \sum_{\sigma_1 \ldots \sigma_\nu \sigma_{\nu+1}}^{k} \sum_{\tau_1 \ldots \tau_\nu}^{n-\rho} M^{\sigma_1 \ldots \sigma_\nu \sigma_{\nu+1}}_{\tau_1 \ldots \tau_\nu n-\rho+1} \frac{\partial (z_{\sigma_1}, \ldots, z_{\sigma_\nu}, z_{\sigma_{\nu+1}})}{\partial (x_{\tau_1}, \ldots, x_{\tau_\nu}, x_{n-\rho+1})} \tag{10}$$

The system (1)—(2) of the $q = (n-\rho)$ class

$$\rho = 1, 2, \ldots, n \tag{11}$$

is on the equation (3)—(7), if through the substitution

$$\left.\begin{aligned} p_{\sigma j} &= Z^j_\sigma - \sum_{i=1}^{q} X^j_i p_{\sigma i} \\ j &= n-\rho+1, \ldots, n \\ \sigma &= 1, 2, \ldots, k \end{aligned}\right\} \tag{12}$$

this equation becomes an identity.

The substitution (12) concerns only the expressions

$$G_{n-\rho+1}, \ldots, G_{n-1}, G_n. \tag{13}$$

If by

$$\{G_{n-\rho+1}\}, \ldots, \{G_{n-1}\}, \{G_n\} \tag{14}$$

are denoted the results of transforming these expressions by means of the substitution (12), we shall have the identity:

$$J_{n-\rho} + \{G_{n-\rho+1}\} + \ldots + \{G_{n-1}\} + \{G_n\} \equiv 0. \tag{15}$$

Hence it follows that the equation (3), linear in Jacobians on which is distributed the system (1) of the $q = (n-\rho)$ class, has the form:

$$(G_{n-\rho+1} + \ldots + G_{n-1} + G_n) - (\{G_{n-\rho+1}\} + \ldots + \{G_{n-1}\} + \{G_n\}) = 0. \tag{16}$$

The symbolic form corresponding to the equation (16) is written down directly.

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PHYSICS

INTENSITY AND POLARIZATION OF RAMAN LINES AND THE FORM OF MOLECULAR VIBRATIONS

By M. ELIASHEVICH and M. WOLKENSTEIN

(Communicated by S. I. Wawilow, Member of the Academy, 3. VIII. 1943)

It has been shown previously by one of us (1) that the intensities and the degrees of polarization of Raman spectra and the intensities of infra-red spectra may be calculated by a semi-empirical method based on the additivity scheme of molecular optics. A further elaboration of these ideas leads to essential additions to the well known rules of Placzek (2). The polarizability of a molecule b_{ik} (where $i, k = x, y, z$) may be expressed in terms of the polarizabilities of individual bonds $\alpha_{n,m}$ (α_{n1}, the polarizability along the bond n; α_{n2} and α_{n3}, along perpendicular directions);

$$b_{ik} = \sum_n b_{ik}^{(n)} = \sum_n \left[\sum_{m=1}^{3} \alpha_{nm} \cos(nmi) \cos(nmk) \right]. \tag{1}$$

The intensity and the depolarization factor of a Raman line are determined by the values of derivatives of the tensor b_{ik} with respect to the corresponding normal coordinate Q_l which is a linear combination of natural vibrational coordinates—of the changes q of bond lengths and the changes γ of valence angles (3). We have

$$\frac{\partial b_{ik}}{\partial Q_l} = \left(\frac{\partial b_{ik}}{\partial Q_l}\right)_\gamma + \left(\frac{\partial b_{ik}}{\partial Q_l}\right)_q \tag{2}$$

where

$$\left(\frac{\partial b_{ik}}{\partial Q_l}\right)_\gamma = \sum_{nt} \frac{\partial b_{ik}^{(n)}}{\partial q_t} \frac{\partial q_t}{\partial Q_l} = \sum_{nlm} \frac{\partial \alpha_{nm}}{\partial q_t} [\cos(nmi)\cos(nmk)] \frac{\partial q_t}{\partial Q_l} \tag{3a}$$

$$\left(\frac{\partial b_{ik}}{\partial Q_l}\right)_q = \sum_{ns} \frac{\partial b_{ik}^{(n)}}{\partial \gamma_s} \frac{\partial \gamma_s}{\partial Q_l} = \sum_{nsm} \alpha_{nm} \frac{\partial}{\partial \gamma_s} [\cos(nmi)\cos(nmk)] \frac{\partial \gamma_s}{\partial Q_l}. \tag{3b}$$

A new physical description of valence and deformation vibrations follows directly from the additivity scheme: according to (3a) and (3b) the intensity of the former depends only on $\frac{\partial \alpha}{\partial q}$, of the latter on the values of α themselves. Let us consider pure deformation vibrations. Their properties are determined by the tensor $\left(\frac{\partial b_{ik}}{\partial Q_l}\right)_q$. The spherical part of such a tensor

$$A_\gamma = \sum_i \left(\frac{\partial b_{ii}}{\partial Q_l}\right)_q = \sum_{nsm} \alpha_{nm} \frac{\partial}{\partial \gamma_s} \left[\sum_i \cos^2(nmi) \right] \frac{\partial \gamma_s}{\partial Q_l} = 0. \tag{4}$$

This may be understood in the following way. The deformation vibrations of polyatomic molecules represent rotations of individual bonds from their equilibrium positions. The spherical part of the polarizability ellipsoid

$$S = \sum_i b_{ii} = \sum_{ik} b_{ii}^{(k)} = \sum_{nm} \alpha_{nm} \left[\sum_i \cos^2 (nmi) \right] = \sum_{nm} \alpha_{nm} \qquad (5)$$

remains invariant because, on the assumption of additivity, the α_{nm} do not depend on bond rotations. Hence the change A_γ of S vanishes. As well known, the depolarization factor of a Raman line is given, in terms of the spherical part $A = \sum_i \frac{\partial b_{ii}}{\partial Q_l}$ and the anisotropic part $B = \sqrt{\frac{1}{2}\left[3\sum_{ik}\left(\frac{\partial b_{ik}}{\partial Q_l}\right)^2 - \left(\sum_i \frac{\partial b_{ii}}{\partial Q_l}\right)^2\right]}$ of the tensor $\frac{\partial b_{ik}}{\partial Q_l}$, by the expression

$$\rho = \frac{6B^2}{5A^2 + 7B^2} \qquad (6)$$

From (4) the following rule is obtained: Raman lines due to pure deformation vibrations have irrespective of their symmetry a depolarization factor 6/7. This rule is an extension of Placzek's ($\rho = 6/7$ for not fully symmetrical vibrations).

Practically, a different ratio of changes of bond length and changes of valence angles may occur for totally symmetrical vibrations, that ratio depending on masses and force constants (except some special cases of such vibrations, when the angles do not change at all). For totally symmetrical valence vibrations A never vanishes. Therefore, all other conditions being the same, the value of ρ may be taken as a characteristics of the part played by the deformation of angles in the vibration considered; ρ may change from small values for pure valence vibrations up to 6/7 for pure deformation vibrations. The characteristic valence vibrations must be polarized, the polarization being greater, the more isotropic the tensor $\frac{\partial \alpha}{\partial q}$ of the bond. The depolarization factor of totally symmetrical vibrations of the same type must be greater for small frequencies due to the greater part played by the deformation in that case.

These considerations are confirmed by experiment (4). Thus, e. g. the Raman line of the symmetrical deformation vibration of the CH_2 group having the frequency ~ 1460 cm^{-1} is always strongly depolarized — ρ is near to 6/7. In the following table the values of ρ are given for characteristic vibrations of the molecules of diatomic type and for totally symmetrical vibrations of molecules of the type XY_2.

From the general formulae (3a) and (3b) some conclusions may be also made with respect to the dependence of these tensors on $\frac{\partial \alpha_{nm}}{\partial q_n}$ and α_{nm}. For pure deformation vibrations, whatever their symmetry, it is easily found that $\left(\frac{\partial b_{ik}}{\partial Q_l}\right)_q$ is a function of the differences $\alpha_{n1} - \alpha_{n2}$ and $\alpha_{n1} - \alpha_{n3}$ for individual bonds, and is for σ-bonds ($\alpha_{n2} = \alpha_{n3}$) a function of the anisotropy $\alpha_{n1} - \alpha_{n2}$ of these bonds. In a similar way $\left(\frac{\partial b_{ik}}{\partial Q_l}\right)_\gamma$ depends, in the case of not totally symmetrical valence vibrations, on the differences $\frac{\partial \alpha_{n1}}{\partial q_n} - \frac{\partial \alpha_{n2}}{\partial q_n}$ and $\frac{\partial \alpha_{n1}}{\partial q_n} - \frac{\partial \alpha_{n3}}{\partial q_n}$ and in the case of σ-bonds simply on $\frac{\partial \alpha_{n1}}{\partial q_n} - \frac{\partial \alpha_{n2}}{\partial q_n}$. Only for totally symmetrical valence vibrations does the dependence on other combinations of $\frac{\partial \alpha}{\partial q}$ appear.

As follows from the types of dependence described, the relative values of the differences $\alpha_{nm} - \alpha_{nm'}$ and $\frac{\partial \alpha_{nm}}{\partial q_n} - \frac{\partial \alpha_{nm'}}{\partial q_n}$ for individual bonds may be obtained

Table of ρ values

Compound	Bond	Frequency cm^{-1}	Type of vibration	ρ
CH_3Cl	C—Cl	712	ν (s)	0.2
CH_3Br	C—Br	594	ν (s)	0.2
C_2H_5Cl	C—Cl	656	ν (s)	0.15
C_2H_5Br	C—Br	560	ν (s)	0.19
C_2H_5I	C—I	500	ν (s)	(0.22)
CH_3SH	C—S	704	ν (s)	0.3
XYCO	C=O	1663—1735	ν	0.34—0.42
CH_2Cl_2	C—Cl	283	δ (s)	0.43
		700	ν (s)	0.09
CH_2Br_2	C—Br	174	δ (s)	0.35
		577	ν (s)	0.11
CH_2I_2	C—I	121	δ (s)	0.42
		483	ν (s)	0.19
$(C_2H_5)_2O$	C—O	438	δ (s)	0.48
		840	ν (s)	0.29
SO_2	S=O	525	δ (s)	0.50
		1145	ν (s)	0.18
CH_2	C—H	~1450	δ (s)	~0.86
		~2850	ν (s)	~0.10

from a comparison of intensities and polarizations of different deformation vibrations and not totally symmetrical valence vibrations.

Let us finally make some remarks concerning the additivity scheme itself. Physically this scheme is well grounded. In particular, the constancy of values α_{nm} for the same bond in different molecules is confirmed by the constancy—caracteristic properties—of electronic frequencies for a given bond, the α_{nm} depending on these frequencies according to the dispersion formula ([5]). The great charges of Kerr constants and depolarization factors of Rayleigh scattering that are observed when we pass to solutions and liquids, are connected with intermolecular interactions—in that case we have to deal not with individual molecules, but with their aggregates. Owing to the independence of vibrations of individual molecules, the considerations concerning the Kerr effect and Rayleigh scattering in gases may as well be applied to the Raman effect in liquids. However, the use of additivity scheme for numerical calculations of intensities in vibrational spectra meets with serious difficulties which are connected with the absence of a reliable mechanical description of vibrations for the overwhelming majority of molecules. The force constants generally referred to in literature have usually little to do with reality, notwithstanding a good agreement between calculated and measured frequencies. The form of vibrations, hence the depolarization factors and the intensities are extremely sensitive to the changes of force constants. In that respect the values of ρ and of the intensities, as obtained in a previous paper ([1]), are to a certain extent fortuitous, because they are based on force constants, that have been determined very roughly. At the present time we have begun a calculation of intensities in spectra of halogen substituted methanes, using a more refined mechanical picture of vibrations ([6]).

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PHYSICS

ROTATION OF POLARIZATION PLANE OF «FORBIDDEN» LINES IN OUTWARD MAGNETIC FIELD

By V. HERMAN

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The rotation of the polarization plane in an external magnetic field in the case of the ordinary (i. e. dipole permitted) resonance scattering has been studied by Weisskopf (1, 2) and, in the light of the classical theory, by many earlier authors.

In the present paper my intention is to discuss this effect in the case of «forbidden» lines, which are lines whose appearance is «forbidden» in dipole approximation, but is allowed in a higher approximation.

As will be shown later, the rotation of the polarization plane of forbidden lines in an external magnetic field is different from the ordinary case. This apparently makes it possible to use this effect for elucidating the nature of forbidden lines, as well as the effect of Zeeman (4) which has so far been the sole criterion of their classification.

Let N_0; M_0 be the initial state of the atomic system under consideration (M_0 denotes the magnetic quantum number; N_0, the rest of the quantum numbers characterizing the state of the system). Accordingly, N', M' characterize the intermediate state; ω_0 is frequency; l_0, polarization; n^0, the direction of propagation of the incident photon; while ω, l and n are the respective symbols for the scattered photon.

By the following expression the intensity of scattering corresponding to the transition N_0, $M_0 \rightarrow N, M$ is determined up to a constant factor depending on frequency, which needs not be here considered (5, 6):

$$J_{NM}^{N_0M_0} \sim \left| \sum \frac{\{Q_{pr}\}_{N'M'}^{N_0M_0} \{Q_{st}\}_{NM}^{N'M'} l_p^0 n_r^0 l_s n_t}{E_{N_0M_0} - E_{N'M'} - i\hbar\Gamma^{N'} + \hbar\omega_0} \right|^2,$$

where $E_{N_0M_0}$, $E_{N'M'}$ are energy levels in the external magnetic field **H**, depending upon magnetic quantum numbers M_0, M'; $\omega_{N_0N} \sim \omega_0$ is the resonance frequency of the forbidden transition; $\Gamma^{N'}$ is the breadth of the level; $\{Q_{pr}\}_{N'M'}^{N_0M_0}$ are the matrix elements of the component of the quadripole moment tensor (the Q_{pr} are so determined that their spur disappears). Here and everywhere below summation is understood to be carried out over the indices that occur twice.

In order to take into account the selection rules of M automatically and to study polarization, the above expression may be conveniently rewritten by using circular coordinates in place of Cartesian: x_3, $\frac{1}{\sqrt{2}}(x_1 + ix_2)$, $\frac{1}{\sqrt{2}}(x_1 - ix_2)$

which we denote respectively by $x_\alpha (\alpha = 0, 1, -1)$ with the unit vectors $\mathbf{L}_\alpha (\alpha = 0, 1, -1)$ equal to $\mathbf{L}_3$; $-\frac{1}{\sqrt{2}}(\mathbf{L}_1 + i\mathbf{L}_2)$, $\frac{1}{\sqrt{2}}(\mathbf{L}_1 - i\mathbf{L}_2)$; the unit vectors $\mathbf{L}_\alpha$ are obviously linked up by the orthogonality relations $\mathbf{L}_\alpha \mathbf{L}_\beta = \delta(\alpha+\beta)$, where

$$\delta(\alpha+\beta) = \begin{cases} 0 & \text{at } \alpha+\beta \neq 0. \\ 1 & \text{at } \alpha+\beta = 0. \end{cases}$$

In these coordinates the expression $J^{N_0 M_0}_{N' M'}$ takes the form:

$$J^{N_0 M_0}_{NM} \sim \left| \sum_{M'} \frac{\{Q_{\alpha\beta}\}^{N_0 M_0}_{N' M'} \{Q_{\rho\tau}\}^{N' M'}_{NM} l^0_{-\alpha} n_{-\beta} l_{-\rho} n_{-\tau}}{E_{N_0 M_0} - E_{N' M'} - i\hbar\Gamma^{N'} + \hbar\omega_0} \right|^2 .$$

$M' = M_0 + \alpha + \beta$ in virtue of the selection rules; $l^0_{-\alpha}$ is the component of $\mathbf{l}^0$ with respect to the unit vector $\mathbf{L}_{-\alpha}$ ($l^0_{-\alpha} = \mathbf{l}^0 \mathbf{L}_{-\alpha}$), correspondingly with $n^0_{-\beta}$, $l^0_{-\rho}$, $n_{-\tau}$; $Q_{\alpha\beta} = \sum \left(x_\alpha x_\beta - \frac{1}{3}\delta(\alpha+\beta) r^2 \right)$.

We will here be mainly concerned with the case when the rotation angle of the plane of polarization in the magnetic field does not depend on the matrix elements $Q_{\alpha\beta}$. Yet, in the general case of $\mathbf{n}^0$, $\mathbf{n}$ and $\mathbf{H}$ arranged arbitrarily in respect to each other and with the atomic system in an arbitrary state, such a dependence is sure to take no place.

This is obvious even if for the reason that, generally speaking, the above cited expression cannot possibly be so represented in the form of two factors, that only one of them would depend upon $Q_{\alpha\beta}$. From this circumstance the discussion of the general case is of little interest. Here we confine ourselves to systems with a non-degenerate initial state.

In order to obtain a rotation of polarization plane typical of forbidden lines the magnetic field $\mathbf{H}$ is defined to be at right angles to the plane $\mathbf{l}^0$, $\mathbf{n}^0$. The vectors $\mathbf{H}$, $\mathbf{l}_0$, $\mathbf{n}_0$ form, then, a right-hand system with the circular unit vectors $\mathbf{L}_0 = \frac{\mathbf{H}}{H}$, $\mathbf{L}_1 = \frac{\mathbf{l}^0 + i\mathbf{n}^0}{\sqrt{2}}$, $\mathbf{L}_{-1} = \frac{\mathbf{l}^0 - i\mathbf{n}^0}{\sqrt{2}}$.

The expression for the intensity assumes now the form:

$$J^{N_0 0}_{N_0 0} \sim \left| \frac{\{Q_{11}\}^{N_0 0}_{N' 2} \{Q_{-1,-1}\}^{N' 2}_{N_0 0} l_{-1} n_{-1}}{\omega_{N_0 N'} + \omega_0 - 2\Delta - i\Gamma^{N'}} - \frac{\{Q_{-1,-1}\}^{N_0 0}_{N',-2} \{Q_{11}\}^{N',-2}_{N_0,0} l_1 n_1}{\omega_{N_0 N'} + \omega_0 + 2\Delta - i\Gamma^{N'}} \right|^2$$

where $2\Delta = 2g\mu_0 H = \frac{eh}{mc} H$ is the splitting typical of the quadripole case and corresponding to the variation of M by as much as ± 2.

If an orthogonal system of $\mathbf{n}$, ϑ, φ is now introduced; where $\mathbf{n}$ has the direction of observation and ϑ and φ are at every point directed toward the side of increasing angles of the spherical coordinates Θ and Φ, we may put $\mathbf{l} = \vartheta$ and, noting that $l_1 n_1 = \frac{\sin\theta\cos\theta}{2} e^{2i\Phi}$,

$$l_{-1} n_{-1} = \frac{\sin\theta\cos\theta}{2} e^{-2i\Phi} \text{ and } \{Q_{11}\}^{N_0 0}_{N' 2} \{Q_{-1,-1}\}^{N' 2}_{N_0 0} = \{Q_{-1,-1}\}^{N_0 0}_{N',-2} \{Q_{11}\}^{N',-2}_{N_0 0}$$

obtain the expression

$$J^{N_0 0}_{N_0 0} \sim \sin^2 2\vartheta \left| 1 - \frac{1}{\rho} e^{4i\Phi + i \operatorname{arctg} \frac{4\Delta\Gamma^{N'}}{(\omega_{N_0 N'} + \omega_0)^2 + \Gamma^{N'} - 4\Delta^2}} \right|^2$$

where ρ is independent of Φ.

Here the rotation angle of the polarization plane in an external magnetic field; i. e. the angle Φ_0 for which the intensity in the presence of a field **H** is a maximum (for $H=0$ the maximum corresponds to $\Phi_0=0$), will evidently be equal to

$$-\operatorname{tg} 4\Phi_0 = \frac{4\Delta\Gamma^{N'}}{(\omega_{N_0N'}+\omega_0)^2+\Gamma^{N'2}-4\Delta^2}. \qquad (1)$$

Thus, the rotation of the polarization plane for forbidden lines as well as for ordinary lines is about **H**, but has a different value and proceeds in the opposite direction. It may be remarked that in the case of forbidden lines, too, if **H** is directed parallel to $\mathbf{n}^0$ so that splitting is equal to $\pm\Delta$, by analogy with the dipole case corresponding to the change of M by ± 1, the angle of rotation is determined by an expression coinciding with Weisskopf's formula ([1]) for the ordinary rotation.

Supposing, indeed, $\mathbf{l}_0 = \frac{1}{\sqrt{2}}(\mathbf{L}_1+\mathbf{L}_{-1})$ and $\mathbf{l}=\mathbf{0}$, we obtain:

$$J_{N_0 0}^{N_0 0} \sim \cos^2 2\theta \left| 1+\frac{1}{\rho}\, e^{i\left(2\Phi-\operatorname{arctg}\frac{2\Delta\Gamma^{N'}}{(\omega_{N_0N'}+\omega)^2-\Delta^2+\Gamma^{N'}}\right)} \right|^2$$

(where ρ does not depend upon Φ), and from this we have for the rotation of the polarization plane:

$$\operatorname{tg} 2\Phi_0 = \frac{2\Delta\Gamma^{N'}}{(\omega_{N_0N'}+\omega_0)^2-\Delta^2+(\Gamma^{N'})^2}. \qquad (2)$$

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PHYSICS

ON THE THEORY OF THE SURFACE TENSION OF METALS

By J. G. DORFMAN

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The superficial properties of metals play a very important rôle in different physical and chemical processes. For this reason it seems interesting to bring a certain clearness into our theoretical ideas concerning this problem.

J. Frenkel proposed in 1917 a theory (1) which considers the surface tension as resulting from the electrostatic action of the outer electronic orbits of the superficial atoms; producing an electric double layer on the surface. This theory, accounting for the action of the «bound» electrons only, applies first of all to non-metallic liquids.

In 1928 J. Frenkel demonstrated (2) that an electric double layer of a rather different kind exists also on metals, owing to a cloud of «free» electrons surrounding the metallic surface. However, the problem considering the surface tension of metals in connexion with the structure and properties of this electronic cloud has not been studied up to the moment, as far as I am aware of it. On the other hand, the existing (2) theory of the potential of the double layer of this electronic cloud permits us to compute the surface tension of metals. The potential $\Phi(x)$ in the space outside the metallic surface ($x>0$) is equal to

$$\Phi(x)=\frac{a}{b^4(c+x)^4 e}, \tag{1}$$

where a and b are coefficients depending on universal constants alone: e, the electronic charge; c, a function of the potential Φ_0 on the metallic boundary ($x=0$). The potential $\Phi(x)$ in the space beneath the boundary (where $x<0$) cannot be calculated at present with a sufficient precision, as has been pointed out by Frenkel and Brillouin (3). Considering the surface energy of the metal as equal to the electrostatic energy of the «cloud» double layer, we obtain

$$\sigma_e=\frac{1}{8\pi}\left(\int_0^{\infty}\left(\frac{\partial\Phi_1}{\partial x}\right)^2 dx-\int_0^{-\infty}\left(\frac{\partial\Phi_2}{\partial x}\right)^2 dx\right), \tag{2}$$

where Φ_1 refers to the space where $x>0$, while Φ_2 refers to the space where $x<0$. It may be easily shown that the rôle of the second term in (2) is negligibly small in comparison with the first, for this reason it may be omitted; thus, we find

$$\sigma_e=\frac{2a^2}{9\pi b^8 c^9 e^2}=4.22\,(\Phi_0)^{\frac{9}{4}}\ \text{dyne/cm}. \tag{3}$$

In a real metal, besides the superficial cloud of the «free» electrons, the «bound» electrons of metallic ions must be taken into consideration. Thus, the «complete» surface tension of the metal

$$\sigma=\sigma_i+\sigma_e,$$

where σ_j is the part of the surface tension due to the influence of the ions. The σ_j may be calculated following the initial Frenkel's theory. In this way

$$\sigma_j = \frac{\pi}{20}(ez_j n)^2 \bar{r}_j^2, \qquad (4)$$

where z_j is the number of the external electrons of the ion; n, the number of the atoms (ions) in a cm^3; r_j, the radius of the external electronic orbit of the ion. This value may be approximately computed from the diamagnetic susceptibility of the ion $\chi_{\partial j}$. We obtain in this way for the surface tension

$$\sigma = \frac{8.7 \times 10^{10} (z_j)^{\frac{1}{2}} (\chi_{\partial j})^{\frac{3}{2}}}{A^2} + 4.22 (\Phi_0)^{\frac{9}{4}}, \qquad (5)$$

where A is the atomic volume.

In case the electronic cloud does not exist, i. e. if the valence electrons of the superficial atoms are «bound» like those in neutral atoms, we find

$$\sigma = \frac{8.7 \times 10^{10} (z_a)^{\frac{1}{2}} (\chi_{\partial a})^{\frac{3}{2}}}{A^2}, \qquad (6)$$

where z_a is the number of the external electrons of the neutral atom, and $\chi_{\partial a}$, its diamagnetic susceptibility.

The formula (5) permits us to compute Φ_0 for normal metals, using experimental data for σ and $\chi_{\partial j}$. The knowledge of Φ_0, the value of the potential at the surface of the metal, permits us further to find the minimum work function φ for the metallic electrons, because

$$\varphi = \Phi_0 - F, \qquad (7)$$

where F is the limiting kinetic energy of the Fermi gas of entirely free electrons. The corresponding values are given in Table 1. The last column but one contains the φ_{calc} computed from (7); in the last column are given the experimental values φ_{expec} ([4]). The values of σ_{exper} in this table are extrapolated to the absolute zero temperature.

Table 1

	σ_{exper}	$\chi_{j\partial} \times 10^6$ exper	z_j	σ_e	Φ_0, volt	F, volt	$\varphi_{calc} = \Phi_0 - F$	φ_{exper}
Na .	223	6 (Na^+)	6.4	216.6	5.76	3.15	2.61	2.25—2.46
Ag .	~1140	26 (Ag^+)	233	~ 900	10.6	5.59	5.0	4.74
Cu .	~1500	17 (Cu^+)	360	~1150	12.0	7.2	4.85	4.82

The data of the Table 1 demonstrate that the work function of the electrons of those monovalent metals may be almost entirely explained by the passage through the double electric layer on the surface of the metal. Thus, the statement of Wigner and Bardeen ([5]) that in monovalent metals, like Na, the rôle of the double layer in the work function may be negligibly small, seems to be wrong.

The examples cited here show clearly that the approximate theory of the surface tension as described by the formula (5) may in a quite satisfactory manner represent the rôle and the properties of the electron cloud at the surface of a metal.

On the other hand, the approximate theory of Frenkel given by the formula (6) represents in an equally satisfactory manner the properties of those liquids, where only «bound» electrons exist on the surface.

Stefan ([6]) has shown on quite general considerations that the surface tension σ is connected with the latent heat of evaporation Q by the following relation:

$$\sigma = \frac{1}{2} \frac{Q}{N^{1/3} A^{2/3}}, \qquad (8)$$

where N is Avogadro's number, and A is atomic volume. As well known; this relation appears to hold for a great number of non-polar liquids whose surface layer

does not possess any special structure. The relation (8) holds in a splendid manner for non-metallic monoatomic liquids as, e. g., the liquefied noble gases Ar and Ne. It seems interesting to investigate, to what extent this Stefan's relation may be applied to metals. Table 2 demonstrates that the formula (8) does not hold for metals of the type of Na, Ag, Cu, Zn, but it represents in a remarkable manner the properties of Hg. For the sake of comparison the noble gases Ar and Ne (liquefied) are also listed in Table 2.

Table 2

	σ_{exper}	σ_{calc} from (8)	σ_{calc} from (6)
Na . . .	223	750	50—60
Ag . . .	1140	2700	300—400
Cu . . .	1500	4650	500—600
Zn . . .	900	1350	300—400
Hg . . .	535	580	500—650
Ar . . .	39	39.5	36.3
Ne . . .	14.7	15.6	15.2

A comparison of σ_{exper} with σ from (8) discloses very distinctly that Hg belongs to a different type of metals than Na, Ag, Cu, Zn. As concerns its superficial properties, Hg stands nearer to non-metallic substances than to normal metals. Apparently, no electronic cloud exists on the surface of mercury; the valence electrons of its superficial atoms seem to be «bound» on discrete levels, much alike those of the orbits in an isolated neutral atom. This circumstance is most forcibly brought out by comparing σ_{exper} with σ calculated by means of (6). When computing σ from (6) we used the most probable values for the diamagnetic susceptibility of neutral atoms. Thus, formula (6) by which the properties of non-metallic substances are represented very satisfactorily, appears applicable to metallic Hg also, but not to Na, Ag, Cu or Zn. We conclude from this state of affairs that metals may be divided into two classes with regard to their surface: a) «normal» metals with a diffuse electron cloud, and b) «abnormal» metals with discrete electronic surface levels.

We may expect that the less «metallic» is the particular metal, the nearer does it stand to the class (b). Apparently, this consideration must apply equally well to metals in a solid state and in the liquid. Thus, for instance, we may conclude from the magnetic and electric properties (7) of the «grey» tin that it belongs to the class (b), while «white» tin positively belongs to (a). A simple computation shows that for this reason the surface energy of grey tin (at low temperatures) must be much below that of white tin. This circumstance makes it easy to understand why white tin is transmitted into grey tin (the so-called «tin-pest») when both phases enter into contact.

This point of view permits also to account in agreement with experience for the observed influence of some surface-active substances when introduced into tin, on the velocity of the reaction in polycrystals (8).

Thus, it may be seen that the theory of surface tension of metals exposed here, though a rough approximation, is none the less in a satisfactory agreement with experimental data and may be used as a point of departure for further investigations of the superficial properties of metals and alloys.

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CHEMISTRY

CRYSTALLOCHEMICAL COMPOUNDS OF SOLID SOLUTIONS. IRON ALLOYS

By N. V. GRUM-GRZIMAYLO

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I have established in an earlier paper ([21]) that the lattice constant of a solid solution depends upon the product of its components' concentrations raised to powers whose exponents are the indices of the chemical formula, each relating to the respective ordered atomic grouping present in the solution.

In order to find out the characteristic features of such crystallochemical compounds I have studied the data secured by different authors on a series of solid solutions in iron. Namely, 1) α-solid solutions: Fe—N, Fe—Cr, Fe—Ni, Fe—Si, Fe—Al, Fe—Mn, Fe—Va, Fe—Co, Fe—Zn, Fe—W, Fe—Cu and Fe—C (tetragonal martensite); 2) ternary α-solid solutions: Fe—Cr—Si and Fe—Mn—Co; and 3) γ-solutions Fe—C and Fe—Pd.

From the lattice constant determinations available in literature it was possible to construct formulae for the lattice constant, and, by the methods indicated in ([21]), uniquely determine the indices in the chemical formulae of the crystallochemical compounds investigated.

The formulae thus obtained are given in Tables 1 and 2.

In that way two main types of crystallochemical compounds in solid solutions have been revealed:

Type I. Crystallochemical compounds in solid solutions of the substitution type. These are superstructural atomic groupings. The main cell of these compounds in a body-centered cubic lattice is formed by one, eight, twenty seven, etc. elementary cells of the centered cube of the α-iron without disturbing the symmetry of the crystal. The formation of such chemical compounds does not depend on the valence of the solute, and the individual properties of the latter only affect the dimensions of the lattice and have no influence on the structure.

In that way the following crystallochemical compounds of type I have been discovered in iron alloys:

a) α-solid solutions FeCr, FeVa, FeCo, FeAl; γ-solid solutions FePd;

b) α-solid solutions Fe_3Cr, Cr_3Fe, Fe_3Ni, Fe_3Si, Fe_3Al, Fe_3Mn, Fe_3Va, Va_3Fe, Fe_3Co and Co_3Fe; γ-solid solutions Fe_3Pd and Pd_3Fe;

c) α-solid solutions $Fe_{25}Cr_2$, $Cr_{25}Fe_2$, $Fe_{25}Ni_2$, $Fe_{25}Si_2$, $Fe_{25}Al_2$, $Fe_{25}Mn_2$ and $Fe_{25}Co_2$;

d) α-solid solutions $Fe_{123}W_3$ and $Fe_{123}Cu_2$;

e) ternary α-solid solutions Fe_8Co_7Mn.

In α-solid solutions all crystallochemical compounds can be built according to one type, by substituting the atom centering the elementary cell of α-iron with the atom of the admixture and arranging four such elementary cells along the diagonals in immediate proximity to one another.

Table 1

Formulae of Lattice Constant for Superstructural Type of Iron Alloy Crystallochemical Compounds

Formulae	Mean quadratic deviation from computed data	Number of determinations	Accuracy of experimental data in decimal places	Authors
γ FePd $\pi=3.5640+0.3187\ Pd+0.121\ FePd+0.1342\ FePd^3+0.5806\ Fe^3Pd$	±0.00027	6	4	Hieltgren, Zapffe [4]
γ FeCr $\pi=2.860+0.0186\ Cr+1.25\ Fe^{25}Cr^2+0.0415\ Fe^3Cr-0.0116\ FeCr+0.007\ FeCr^3-1.0\ Fe^3Cr^{25}$	0.00036	29	4	Preston [5]
α FeNi $\pi=2.86074-0.0669\ Ni-1.743\ Fe^{25}Ni^2+0.1667\ Fe^3Ni$	0.000075	4	5	Jette, Frank-Foot [6]
α FeSi $\pi=2.8604-0.228\ Si+2.8\ Fe^{25}Si^2+0.165\ Fe^3Si$	0.00005	16	4	Jette, Greiner [7]
α FeAl from 0.0—18.5 at.% Al $\pi=2.8613+0.249\ Al+2.7\ Fe^{25}Al^2-0.177\ Fe^3Al$	0.00032	5	4	Bradley, Jay [8]
α FeAl from 18.5—25 at.% Al $\pi=2.8944-0.0155\ Fe_{free}$ where $Fe_{free}=1-4\ Al$	0.00005	4	4	Do
α FeAl from 27.88—51.4 at.% Al $\pi=2.9026-0.0081\ Fe_{free}+0.177\ Fe^2{}_{free}\ Al+0.1787\ Fe^3{}_{free}$ where $Fe_{free}=1-2Al$	0.00040	7	4	Do
α FeMn $\pi=2.857-0.2\ Mn-10.0\ Fe^{25}Mn^2+0.52\ Fe^3Mn$	0.0010	5	3	Friauf [9]
α FeCrSi $\pi=2.8591+[0.0186\ Cr+1.25\ Fe^{25}Cr^2+0.0415\ Fe^3Cr-0.0116\ FeCr]-[0.228\ Si-2.8\ Fe^{25}Si^2-0.165\ Fe^3Si]$	0.00056	34	4	Anderson, Jette [10]
α FeV $\pi=2.86+0.18\ V+0.078\ Fe^3V-0.25\ FeV+0.282\ FeV^3$	0.0068	12	3	Osawa, Oya [11]
α FeV $\pi=2.86+0.18\ V+0.217\ Fe^3V-0.3\ FeV+0.066\ FeV^3$	0.0053	11	3	Wewer, Jellinghaus [12]
α FeCo $\pi=2.857-0.191\ Co-0.037\ Fe^3Co+0.2\ FeCo+0.61\ FeCo^3$	0.00157	113	3	Osawa, Oya [13]
α FeCo $\pi=2.8607-0.058\ Co+1.784\ Fe^{25}Co^2+0.055\ Fe^3Co+0.0375\ FeCo-0.078\ FeCo^3$	0.0000	13	4	Ellis, Greiner [14]
α FeCoMn $\pi=2.861-0.191\ Co+0.037\ Fe^3Co+0.2\ FeCo+0.61\ FeCo^3-0.2\ Mn-10.0\ Fe^{25}Mn^2-0.52\ Fe^3Mn^3+1000\ Fe^3MnCo^7$	0.002	24	3	Köster, Schmidt [15]
α FeCu $\pi=2.8607+55.0\ Fe^{123}Cu^3$	0.00011	6	4	Norton [17]
α FeW $\pi=2.8607+0.25\ W-123.0\ Fe^{123}W^2$	0.00002	4	4	Chartkoff, Sykes [16]

Then the admixture atoms will form a tetrahedron with side as long as the diagonal of the elementary cell of α-iron. The existence of such structures has been proved in Fe_3Si and Fe_3Al alloys. In compounds of the $Fe_{25}R_2$ type the elementary tetrahedron is on every side surrounded by two layers of iron atoms, and in the case of the crystallochemical compound $Fe_{123}R_2$, by seven layers. As is seen from Table 1, all solid solutions of that type have the same formula for the lattice constant:

$$\pi = \pi_0 \pm A_1 C_R \pm A_2 C_{Fe}^{25} C_R^2 \pm A_3 C_{Fe}^3 C_R \pm$$
$$\pm A_4 C_{Fe} C_R \pm A_5 C_{Fe} C_R^3 \pm A_6 C_{Fe}^2 C_R^{25}.$$

In the case of limited solubility this series breaks at the respective crystallochemical compound comprised within the region of the solid solution. An analogous formula has been established for solid solutions of γ-iron. For alloys with very limited region of solid solutions the formula of the constant was found to be

$$\pi = \pi_0 \pm A_1 C_R \pm A_2 C_{Fe}^{123} C_R^2.$$

Table 2

Formulae of Lattice Constant in Iron Alloys (Crystallochemical Compounds of Molecular Type)

Formulae	Mean quadratic deviation from computed data	Number of determinations	Accuracy of exper. data in decimal places	Authors
γ FeC: $\pi_{800^\circ} = 3.6425 + 4.3386\ C - 1.196 \cdot 10^6 \cdot Fe^{18}C^6 - 4.3943\ Fe^3C$	0.00023	6	4	Esser, Müller (1)
γ FeC: $\pi_{1100^\circ} = 3.6600 + 5.489\ C - 1.322 \cdot 10^6 \cdot Fe^{18}C^6 - 5.3776\ Fe^3C$	0.00035	5	4	Do
αFe—N $\pi = 2.8607 - 0.3746\ N + 131.5\ Fe^{32}N^2$	0.00008	3	4	Eisenhut, Kaupp (18).
αFe—Zn $\pi = 2.86 - 0.006\ Zn + 5.65\ FeZn^3$	0.0006	5	4	Osawa, Ogawa (19)
Tetragonal martensite				
Axis $C\ \pi_C = 2.8610 - 9.42\ C + 793.0\ Fe^{24}C^2 + 10.0\ FeC^5$	0.00032	7	4	Hägg (2), Ohmann (3)
Axis $A\ \pi_A = 2.8610 - 22.0\ Fe^{24}C^2 + 4.0000\ FeC^5$*	0.00027	7	4	Do

Note. Ohmann's data point to the possibility of discarding the third term in the formula for π_A.

Type II. To the second type of crystallochemical compounds in solid solutions belong those formed by atoms that are placed mainly outside the nodes of the lattice, thus lying in the interatomic spaces. The chemical formulae of these compounds are more difficult to determine, because the law of symmetry is here not so rigorously observed as in the previous case. In determining their chemical formulae we have to solve exponential equations involving more than one unknown quantity, and then select the suitable solutions taking other properties of the alloy into consideration.

In the case of ferrocarbon alloys, for instance, I started from the following formula of the lattice constant:

$$\pi = \pi_0 + A_1 C_C + A_2 C_{Fe}^m C_C^n + A_3 C_{Fe}^3 C_C.$$

Substituting in it the experimental data obtained by Esser and Müller I determined the crystallochemical compounds FeC_6 and $Fe_{18}C_6$.

The former of these may be represented by analogy with diamond. The molecule within the crystalline structure of the diamond is a body-centered tetrahedron

formed by five carbon atoms. Such a diamond molecule may possibly add to itself one atom of iron, and then we have our hypothetical compound. The second compound $Fe_{18}C_6$, a polymer of cementite Fe_3C, may be conceived as a graphite molecule formed by six atoms of carbon to which 18 atoms of Fe are added.

I also made an attempt to find the formula of the parameter in tetragonal martensite. The latter was found to contain two crystallochemical compounds: one of them, $Fe_{24}C_2$, is responsible for the tetragonal distortion of the cubic lattice, and the other, FeC_5, may perhaps account for the hardness of martensite.

A crystallochemical compound of the second type was also discovered in the alloys of iron and nitrogen. Its chemical formula is $Fe_{32}N_2$.

Iron-zinc solid solution likewise contains a crystallochemical compound, $FeZn_3$. It has been referred to the second group on the sole consideration that its elementary composition does not fit within the boundaries of a solid solution of zinc in iron.

So in point of structure the second type of compound is to be considered as molecular grouping of atoms lying within the lattice, while the lattice atoms are partly involved in the making of the molecule.

To sum up, the crystallochemical compound of a solid solution presents an ordered distribution of the admixture atoms within the crystalline lattice of the main component, these atoms being held in their respective positions at the nodes or in the interatomic spaces by the forces of valence and crystallization.

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CHEMICAL TECHNOLOGY

PRODUCTION OF CLINKERLESS CEMENT WITH UTILIZATION OF HEAT FROM MOLTEN BLAST-FURNACE SLAGS FOR CALCINING OF ACTIVATING ADMIXTURES

By P. P. BUDNIKOV, Corresponding Member of the Academy

In manufacturing slag clinkerless cement some fuel is expended of necessity on calcining the blast-furnace slag activators, gypsum for anhydrite at 600—700° and dolomite at 800—900° C in the case of basic granulated slags, and at 1000—1100° C with acid slags (1). Searching for further means by which the production of clinkerless cement might be made a more simple and less expensive procedure, I have made use of the heat of molten blast-furnace slags in calcining the above-mentioned admixtures. Gypsum, when introduced into molten blast-furnace slags at about 1400° C, gives off its hydrate water and is partly decomposed with separation of SO_3; the calcium oxide thus formed is dissolved in the metasilicate of the slag (2). The processes that take place on the introduction of gypsum into molten slags are expressed schematically by the following equations:

$$CaSO_4 \cdot 2H_2O = CaSO_4 + 2H_2O; \quad 2CaSO_4 = 2CaO + 2SO_2 + O_2;$$
$$3CaSO_4 + CaS = 4CaO + 4SO_2 + 2O_2; \quad 3CaSO_4 + CaS + 4CaSiO_3 = 4Ca_2SiO_4 + 4SO_2.$$

The anhydrite which is left undecomposed playshere the rôle of the slag activator (3).

On this basis I started industrial experiments with admixtures introduced into the molten slag, while it is admitted into the ladle. A series of such tests has shown that the heat of the molten slag may be turned to profitable account in calcining the admixtures at the time when the ladle is filled with slag up to $^3/_4$ of its height. The admixtures brought into contact with a portion of molten slags in the ladle are calcined and remain on the surface of the latter. The fact that the admixtures come floating on the surface together with a portion of the slag makes it easy to separate the cooled crust from the molten mass after the slag has been poured into a granulator. This crust upon cooling is used as activator of granulated blast-furnace slags in the manufacture of clinkerless cement and consists mainly of anhydrite, calcined dolomite and slag. The slags experimented upon were of the following chemical composition (in per cent).

	SiO_2	Al_2O_3	CaO	MgO	MnO	FeO	SO_3
Foundry slag	33.04	9.00	52.31	0.88	1.11	0.99	2.74
Bessemer slag	35.36	7.40	48.30	2.70	1.42	1.42	3.29
Open-hearth slag	37.99	8.18	45.33	4.37	2.07	0.58	2.39

Gypsum and dolomite were calcined at the expense of the heat evolved by the molten slags at the time when the ladle up to $^3/_4$ of its height was filled with slag. They were taken in the following proportions:

I . . . . . 0.5 gypsum + 0.5 dolomite + 1 molten slag
II . . . . . 0.5 » + 0.5 » + 1.5 » »
III . . . . . 0.5 » + 0.5 » + 2 » »

The investigation of the cements (standard specimens) obtained on the basis of Bessemer slag (chemical composition see above) has furnished the data referred to in Table 1.

Table 1

Composition of cement	Setting time	Grade of grinding (residue on sieve in %)		Ultimate strength in kg/cm² (1 : 3)							
				Tensile				Compression			
		900 mesh per cm²	4900 mesh per cm²	4 days	7 days	28 days	2 months	4 days	7 days	28 days	2 months
70 % of Bessemer slag +30 % of mixture I	9 hrs 40 m.	0.50	7.5	3.5	7.4	10.5	13.4	15	27	49	63
70 % of Bessemer slag +30 % of mixture II	5 » 17 »	0.45	7.1	4.9	12.3	17.3	21.8	72	89	150	183
70 % of Bessemer slag +30 % of mixture III . . .	2 » 05 »	0.50	7.7	6.7	15.8	22.9	23.9	91	107	205	214

From Table 1 it appears that the most active is mixture III; composed by 0.5 parts of gypsum, 0.5 parts of dolomite and 2 parts of molten slag.

The data on the influence mixture III has on typical basic granulated blast-furnace slags (composition see above) added, while ground, with 5 per cent bi-aqueous gypsum, are listed in Table 2.

The best results, as seen from Table 2, were obtained with foundry and open-hearth slags.

Table 2

Composition of cement	Setting time	Grade of grinding (residue on sieve in %)		Ultimate strength in kg/cm² (1 : 3)							
				Tensile				Compression			
		900 mesh per cm²	4900 mesh per cm²	4 days	7 days	28 days	2 months	4 days	7 days	28 days	2 months
65 % of foundry slag +30 % of mixture III+5 % of gypsum	1 hr 45 m.	0.6	6.9	9.1	17.3	24.3	26.8	101	115	217	234
65 % of Bessemer slag +30 % of mixture III+ 5 % of gypsum	2 » 20 »	0.5	7.7	7.2	16.0	23.2	24.4	90	110	207	238
65 % of open-hearth slag+30 % of mixture III+5 % of gypsum	3 » 45 »	0.5	7.0	4.3	12.9	17.2	20.1	79	92	185	195

The clinkerless cement obtained without fuel expenditure on calcining the activating admixtures was tested for the length of keeping and for resistance to the action of mineralized water. 65 per cent of the Bessemer granulated slag (composition see above) were taken for these tests, and while grinding it 30 per cent of mixture III and 5 per cent of gypsum were added. The results of these tests are given in Table 3.

Table 3

Duration of keeping	Setting time	Ultimate strength in kg/cm² (1 : 3)							
		Tensile				Compression			
		Fresh water		Mineralized		Fresh water		Mineralized*	
		7 days	28 days	7 days	28 days	7 days	28 days	7 days	28 days
10 days	2 hrs 20 m.	16.0	23.2	15.8	23.5	110	207	112	204
2 months	2 » 05 »	15.7	24.1	16.1	23.0	108	210	113	201
3 »	3 » 55 »	14.1	21.0	14.7	20.2	99	182	103	189
6 »	7 » 10 »	12.4	17.9	12.0	18.5	83	171	89	183

* The composition of mineralized water, according to A. A. Baikov: 400 g $MgCl_2$, 100 g $MgSO_4$, 80 g NaCl per 1 liter of water. The standard samples were immersed in mineralized water 4 days after their preparation.

During the first two months of storage neither the terms of setting nor the strength of the cement of the above-cited composition proved to be liable to any appreciable alteration of composition. Only after three months had elapsed; was there observed a retardation of setting and a reduction of strength.

Conclusions. The production of clinkerless slag cement is simplified and made cheaper by carrying on the calcining of activator admixtures (gypsum and dolomite) at the expense of the heat available in molten blast-furnace slags. This may be attained by the following methods: 1) the activators are introduced in the flow of molten slag before it is admitted into a dry granulator, and afterwards the slags are ground to cement which is then ready for use [3]; 2) the activators are introduced into the slag ladle after $^3/_4$ of it are filled with molten slag, and this is followed by the separation of the floating admixtures for which purpose the slags are poured into a dry or wet granulator. The calcined admixtures separated from the slag are to be used as activating mixture in the manufacture of clinkerless cement.

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GEOCHEMISTRY

RESULTS OF SPECTRUM ANALYSIS OF ANHYDRITE FROM THE PERMIAN DEPOSITS OF TATARIA

By L. M. MIROPOLSKY and S. A. BOROVICK

(Communicated by V. G. Chlopin, Member of the Academy, 4.V.1943)

Anhydrite is a mineral typomorphic for the Permian deposits of Tataria. According to its distribution it occupies the one of the foremost places. Particularly extensive is the development of anhydrite among the Lower Permian deposits. In the Upper Permian strata segregations of this mineral may be traced much less frequently and only as relicts amidst gypsum. This fact points to a still wider development of anhydrite over the territory of Tataria in the past. Now large accumulations of anhydrite are confined only to Lower Permian deposits, where it forms both small isolated concretions and thick occurrences ([1]; [2]).

Spectrum investigations of anhydrite were made with the purpose to establish with greater precision its chemical composition for different stratigraphical horizons of the Permian deposits according to the best studied deposits, to find out the extent of distribution of individual elements. Five specimens of anhydrite were subjected to qualitative spectrum analysis: two from the region of the Sokolyi Gory village, one from the Kamskie Polyany village; one from the Kotlovka village and one from Kamskoye Ustye.

The spectrum photographs were taken under the following conditions: the material was burned in the flame of an a. c. voltaic arc 220 volts, 10 amperes; the photographs of this flame were taken with a Zeiss quartz spectrograph Q-24. Of each specimen two spectrographs were successively taken; the second of these would give intenser lines of poorly volatile elements and ionized atom lines. The anhydrite specimens were analysed for the following fifty six elements: Li; Be, B, Na; Mg; Al; Si; K; Ca, Sc, Ti; V; Cr, Mn; Fe; Co, Ni; Cu, Zn, Ga, Ge, As; Sr, Y, Zr, Nb; Mo; Ag, Cd, In; Sn, Sb; Te, Ba; La; Ce; Pr, Nd, Il; Sm; Eu, Gd, Tb, Dy, Ho, Er, Tu, Yb, Cp, W, Os, Ir, Pf, Ta, Pb, and Bi. The presence of only ten of these elements was established in the anhydrite: Mg, Al, Si, Ca, Ti, Mn, Fe, Cu; Sr and Zr.

The place of sampling of anhydrite; the morphological peculiarities of the specimens and the results of the interpretation of the spectrographs are tabulated below (page 383).

The examination of the table permits of the following conclusions:

1) All anhydrite specimens have a rather simple chemical composition; accessory elements are few, and most of them show a quite insignificant concentration.

2) The anhydrite specimens show no sharp distinction in composition for different stratigraphical horizons; their elementary composition is almost coincident except for the anhydrite from the Upper Permian deposits, which contains Cu and Zr.

Place of sampling and its principal properties	Age of deposition	Elementary composition									
		Mg	Al	Si	Ca	Ti	Mn	Fe	Cu	Sr	Zr
Sokolyi Gory vill., right bank of the Kama, hole No. 6. Anhydrite fibrous, blue	P_1^A	b	1	b	A	1	1	1	—	b	—
Same, hole No. 6. Anhydrite fibrous, light-blue	P_1^k	b	2	b	A	1	1	1	—	b	—
Kamskie Polyany vill., left bank of the Kama, hole No. 101. Anhydrite bluish, of granular-fibrous structure	»	1	2	b	A	1	4	1	—	a	—
Kotlovka vill., right bank of the Kama, hole No. 201. Anhydrite fibrous, blue-coloured	»	1	1	1	A	2	3	2	—	b	—
Kamskoye Ustye vill., right bank of the Volga, at the «Suvod» ravine, near the «9th of January» Works, from the «semisagennuy» layer. Anhydrite bluish, of granular fibrous structure	$P_2^{kaz_3}$	2	3	1	A	a	—	1	3	a	3

Denotation of the brightness of lines: A, extremely intense brightness; B, very intense; C, intense; a, above medium; b, medium; 1, weak; 2, very weak; 3, traces; 4, negligible traces; 5, hardly discernible.

3) In the anhydrite specimens leading part always belongs to Ca; they show high concentration of Mg, Sr, Si, sometimes of Ti, Al, Mn, Fe and only traces of Cu and Zr.

4) The anhydrites from the Lower Permian deposits are quite similar as to their qualitative elementary composition, while the anhydrites from Upper Permian deposits contain an additional admixture of Cu and Zr.

5) The content of Ti and Mn fluctuates within a rather considerable range when evaluated by the intensity of the lines, while the content of the rest of elements, as judged by the intensity of the corresponding lines, shows only negligible discrepancies and deviations.

6) According to the frequency of distribution, the elements may be classed into three groups: 1) elements invariably present—Mg, Al, Si, Ca, Ti, Fe, Sr; 2) elements of frequent occurrence—Mn, and 3) elements sporadically observed—Cu and Zr.

When comparing the results of the qualitative spectrum analysis of the specimens of anhydrite, celestine, and gypsum previously described [3] it may be seen, first, that the composition of the chemical components is similar; secondly, that their dissolution is of the same order; and thirdly, that the composition of anhydrite is even simpler.

The invariable presence of Mg, Al, Si, Fe and Sr in all the anhydrite specimens makes us assume that a certain amount of them is present in gypsum, probably as a mechanical admixture in the form of dolomite, quartz, chalcedony, opal, goethite, hydrogoethite, celestine and finely silty material. The greater part of the elements enter into the lattice of anhydrite as isomorphic admixtures, following the principle of compensational and polarly directed isomorphism.

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GEOLOGY

THE IMPORTANCE OF FORMS OF GRANITE PEGMATITES

By C. VLASSOV

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In the study of deposits of rare and disseminated metals associated with granite pegmatites (the Altai, the Urals, Transbaikal, North America, South America, Africa), attention is drawn to the following facts:

I. It is only an insignificant part of the total large number of pegmatite fields of the world that are characterized by a large amount of rare-metal minerals, and only a very small number of veins from a few fields contain accumulations of such minerals in commercial quantities. Most of these veins carry rare-metal minerals in very small amounts, or even practically do not contain them at all.

II. The largest and most concentrated segregations of minerals of lithium, caesium, rubidium, beryllium, niobium, tantalum, bismuth, tin and other elements are found in the largest pegmatite bodies, particularly in those having an oval shape (lenses, stocks, etc.).

III. As a rule, the minerals of the above-named elements occur in pegmatites as nests, which in large pegmatite bodies are confined chiefly to the central parts, while in veins of small width (up to a few meters) this rule does not hold.

IV. The zoning, i.e. the differentiation into separate mineralogical complexes; and in some cases monomineral zones too, is expressed better in bodies of an oval shape and of greater width, than in those of a platy form and insignificant width; but of the same size. The largest pegmatite bodies, particularly those of an oval shape, are characterized by the largest mineral individuals.

In many cases in pegmatites of an oval shape a concentric arrangement of zones is observed. The zones are generally distributed as follows: the outer «ring» consists of a pegmatite with a graphic structure, up to some meters in width; this is followed by a «ring» of practically monomineral microcline, sometimes attaining a width of several tens of meters; this in turn is followed by a central part consisting of blocks of quartz, and of the following minerals: spodumene, beryl (sometimes rich in caesium), columbite, cassiterite, apatite, lepidolite, albite (clevelandite) formed to a large extent at the expense of microcline (often containing rubidium), muscovite, polychromic tourmalines, garnets rich in manganese, well-formed crystals of quartz, etc. In some cases the central ore complex mentioned is subdivided into a number of complexes or independent zones: the albite-spodumene, the beryl-muscovite-garnet, and other zones.

The ore complexes described are usually late in time of formation, their development being accompanied by a replacement of different earlier minerals or their complexes; as a rule, they are confined to the large-block parts of pegmatites, made up chiefly by coarsely crystalline microcline and quartz blocks.

V. The replacement processes are best manifested also in large pegmatite bodies, particularly in those of an oval shape.

VI. In platy pegmatite bodies of small width (up to a few meters) the zoning is displayed much less clearly. The ore nests are either absent altogether, in which case the rare-metal minerals are isolated from one another and distributed throughout the whole body of the veins, without any evidence of localization, or else these nests are small and scattered at random, sometimes occupying the periphery of the veins. In wide plate-shaped bodies (over 10 m in width) the zoning is displayed better than in those of insignificant width. In the latter the ore bodies are found mainly in the central part, but they are also smaller than in oval-shaped bodies of the same size. The volume of the minerals is smaller than in lens-shaped bodies.

VII. When analysing the data available, three questions arise, namely:

1. Why do the pegmatite bodies belonging to the same pegmatite field and having about the same size and chemical composition, but different shapes, differ so markedly in their structure and particularly in the content of rare-metal minerals?

2. Why do the ore nests have a larger size in oval-shaped bodies, and why are their minerals larger?

3. Why does the ore-forming process in pegmatites of an oval shape usually develop in the central parts, the replacement process being intenser there?

From his own field observations of a number of pegmatite fields, as well as from an analysis of the Russian and foreign literature on granite pegmatites (particularly those carrying economic amounts of rare-metal minerals, such as spodumene, lepidolite, beryl, tantalite, cassiterite, minerals of bismuth, polychromic tourmaline, etc.), the author is enabled to make the following conclusions:

In the process of mineral formation of the granite-pegmatite melt in the place of its crystallization, i. e. in the country rock, a highly important part is played by the shape assumed by the pegmatite melt.

The shape that is advantageous for the process of mineral formation and for the attendant concentration of rare elements and localization of rare-metal minerals, is the spherical one, or that approaching the sphere, such as lenses and compact bodies of other shapes. The least favourable forms are narrow veins with parallel walls. An intermediate position is occupied by wide veins or veins with swells.

The ratio of external surface to volume being lower in oval bodies, their cooling area is smaller as compared with that of bodies of a platy shape of the same size. Bearing in mind the usual natural sizes of pegmatite bodies, it can be said that the external surface of a platy body is forty-sixty times that of a spherical one.

The relatively small cooling area of oval bodies is responsible for the rate of temperature decrease which is a very low one, owing to which the process of crystallization takes a long time; in other words, there is a very slow process of mineral formation. The considerable period of mineral formation results in a more distinct manifestation and separation of the stages of the pegmatite process, and consequently, a complete spatial and physical-chemical differentiation of matter. The pegmatite melt begins to crystallize from the periphery. This phenomenon which is related with the concentration of the components and with temperature determines in principle also the spatial distribution of the minerals; on the periphery are deposited such minerals whose formation temperature is the highest for the minerals of a given body, while low-temperature minerals are deposited in the centre. As the rare elements have no sufficient concentration at an early stage of crystallization of the pegmatite melt and thus cannot at this stage be considered as independent components, they, as a rule, crystallize at a later stage as compared with the principal minerals, inclusive of microcline and a considerable part of quartz. Owing to this, in the course of crystallization they are partly pressed towards the central parts by the feldspar and quartz zones. A part of them—the greater, as it seems,—always remain in the minerals of the earlier generations in a disseminated condition. This is evidenced by the fact that

a number of rare elements are found in feldspars and other minerals of earlier formation. In platy bodies of small width the process of crystallization goes on at a higher rate, therefore, the differentiation of matter is less complete, and a considerably larger part of the rare elements become disseminated.

Moreover; in platy bodies the squeezing out of rare-metal compounds at the time of crystallization proceeds practically from two sides—from the side walls towards the central portion which is more or less parallel to the walls. In oval bodies the squeezing goes on from all sides towards the centre, and thus in the inner parts of the lenses rare elements are accumulating, which are as it were being extracted from the enormous volume of the pegmatite melt. In platy bodies of a similar size there is no such possibility, since in this case the ore solutions are pressed towards the cavities of the pegmatite bodies, their lateral solid surfaces being many times those of the oval bodies. That is why ore solutions cannot form large accumulations in one place.

In platy bodies pinchings are observed both along the strike and along the dip. The vein is thus divided during crystallization into a system of practically isolated sections with independent foci of ore solutions. This is not the only cause that prevents large-scale localization of ore solutions in platy bodies. It should be also borne in mind that prior to crystallization, ore solutions usually react with zones of minerals of earlier generations, such as microcline and some others. In veins of a platy shape the reaction areas are many times larger than, e. g.; in lenses; i. e. these solutions have a larger contact and reaction surface with the earlier zones. Thus, the ore solutions become again less concentrated at the expense of the decomposition of the early minerals and an assimilation of their components, and are distributed throughout a greater portion of a given pegmatite body. No large-scale localization can take place even in the case if the solutions would tend to migrate in one direction, say, upwards. In this case, too, the angles of the position of the bodies being also considered, the solutions are localized in a large number of places, i. e. at different levels along the dip and in different places along the strike, no large ore accumulations being formed.

In the process of formation of some mineral, along with such factors as temperature, pressure, etc., a very important part is played by the concentration of the components of a given mineral (law of active masses). If we take into consideration that, being two opposite sides of the same process (law of unity of concentration and dissemination), concentration and dissemination of chemical elements assume a different course in pegmatites of different shape, it is therefore but natural to conclude that in the portions of the pegmatite granite melt which are of similar size and identical in their initial chemical composition; but have acquired a different shape prior to crystallization, the mineral-forming processes will assume a different course. This will result in a different mineralogical composition and in the development of different types of pegmatites. In some cases rare-element compounds may tend towards dissemination; this may so much prevail over the concentration that the rare elements will not be able to form independent components. Thus, in a number of cases there occurs no concentration sufficient to form independent rare-metal minerals: spodumene, lepidolite, cassiterite, beryl, etc. The rare elements enter the lattice of other minerals (such as lithium and tin-bearing muscovites, rubidium and caesium feldspars, etc). In other cases a concentration takes place which proceeds up to the stage when rare-metal minerals are formed; in some cases these are localized even in workable amounts.

The part played by shape and size in the process of localization of rare-metal minerals in granite pegmatites may also account for the fact that the most advantageous enclosing rocks for large deposits of rare-metal minerals associated with pegmatites are gabbros; ultrabasites, and other non-schistose rocks. This is due to the fact that in such rocks fissures with parallel walls are less frequently formed; consequently, the pegmatites occurring in them assume an oval or more compact shape more frequently than; e. g.; in schists.

From field observations it may also be seen that most favourable for the localization of rare-metal minerals are pegmatites which though belonging to the platy

type, have angles of dip ranging between 45 and 90°. Flat pegmatite veins are unfavourable for the formation of workable concentrations of rare elements.

This phenomenon may be accounted for by the fact that with steeper angles of dip of the veins; better conditions for localization are created, volatile and readily fusible compounds tending to ascend. Consequently, the areas of localization are fed from larger columns of pegmatite melts than are bodies with a horizontal or flat attitude.

Both data available and theoretical investigations on granite pegmatites testify of their being an essentially closed system, that is, since the moment of the intrusion of the pegmatite melt into the enclosing rocks, the relation between the pegmatite melt and the parent intrusion practically ceases.

All the phenomena of replacement, filling of fissures, migration and localization of late ore-forming solutions, take place within the pegmatites themselves and at the expense of a portion of the same pegmatite melt. The chemism of replacement is satisfactorily explained from the standpoint of Bowen's reaction principle.

Thus, it becomes clear that the process of formation of economic deposits of rare elements in granite pegmatites is determined by two main factors: first, by the degree of «maturing» of the pegmatite melt in the body of the parent intrusion; here a significant part is played by the quantity of feldspar and quartz which will enter as granite before the pegmatite melt intrudes the enclosing rocks. Secondly, the process of accumulation of rare-metal elements and minerals in pegmatites is determined by the nature of evolution of the melt in the enclosing rocks. In this evolution a very important part belongs to the shape and size of the pegmatite bodies.

As all ore-bearing solutions are volatile and fusible distillates of magmatic intrusions, it is evident that in the course of formation and localization of ore-forming solutions an essential part is played by the shape and size, as well as the mode of occurrence of the parent magmatic bodies. Thus, all the specific peculiarities being considered, observations of pegmatites may be extended to other magmatic phenomena. Consequently, a thorough study of the shape and mode of occurrence of pegmatite bodies and intrusions is highly important in the search, prospecting and working of various deposits.

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GEOLOGY

THE COALS OF THE TAS-KOMYRSAY DEPOSIT

By B. I. KUSTOV

(Communicated by V. G. Chlopin, Member of the Academy, 1. VI. 1943)

The northernmost deposit of the Karatau coal-bearing belt stretching 150—170 km along the separate spurs of the Karatau range, is the Tas-Komyrsay deposit located in the Suzak region of the South Kazakhstan District (Kazakh SSR). The deposit is within 50—55 km south-east of the polymetallic Achisay mine. The deposit known since 1869 was in part prospected in 1927 and 1932—33 (1), and has been subjected to detailed prospecting since 1939 by the Kazakh Geological Service.

The area of the deposit is composed of Jurassic deposits which from their lithological features may be subdivided into four series: the conglomerate, conglomerate-sandstone, the coal-bearing, and the shale series. The coal-bearing series of Lower Jurassic age has as yet no complete section, and its thickness has not been established. The prospected portion of the coal-bearing series contains ten coal seams of which five have been subject to detailed prospecting, and their reserves have been estimated. The thickness of these seams (1^{H}, 1^{b}, 2, 2^{1}, 3) ranges between 0.6 and 2.5 m.

In 1938 the «Achpolymetal» combinate sunk in the deposit a vertical shaft 37 m deep, which at present is not worked because of the conditions of coal transportation.

The coals of Tas-Komyrsay are black, lustrous, with a conchoidal fracture; they withstand weathering and prolonged storage. The coal burns with a bright long flame, almost without soot.

The results of a technical analysis of coals from the coal seams prospected are tabulated in Table 1. When examining the table it should be borne in mind that the ash content of most coal samples is somewhat increased at the expense of bands of coaly and clayey shales, while the ash content of pure coal is only 12 to 15 per cent.

From the content of volatile matter the coals may be referred to the long-flame coals of Grüner's classification. A study of the distributions of sulphur forms in the coals of seam I has shown that pyritic sulphur constitutes on the average 47.2 per cent of the total sulphur of the coal; and organic sulphur, 45.7 per cent of the total sulphur. The elementary composition of the coals of seam I shown in Table 2 is also characteristic of the coals of the I class of Grüner's classification, both with regard to the carbon and oxygen content, and to the $\frac{O+N}{H}$ ratio.

Thus, by their chemical composition the Tas-Komyrsay coals may be referred to coals of the intermediate type between brown coal and stone coal, more closely approaching the latter.

Table 1

Seam	Samples	W^p	A^c	V^r	S^c total	Q_b^g cal/kg
1^H	from holes	7.1—10.3	13.8—30.6	40.8—45.8	0.33—3.6	7040—7580
1^H	average	8.55	21.7	43.13	1.76	7350
1^b	from holes and ore plot of the shaft	6.7—10.0	12.75—26.15	40.4—45.37	0.34—1.60	7460—8030
1^b	average	8.3	21.25	42.81	1.6	7660
1^b	average from the shaft	8.24	25.70	42.98	0.72	7228
2	from the shaft	7.65—8.4	18.0—18.14	40.6—40.9	0.59—0.8	7200—7620
2^1	do	6.3—9.02	10.3—35.8	39.08—40.9	0.64—3.5	7380—7410
2 and 2^1	average	8.18	21.25	40.62	1.66	7410
3	from the shaft	7.97—9.7	8.84—26.0	39.8—45.14	0.39—2.10	7210—7450
3	average	8.96	20.28	39.7	1.53	7285

The average composition of the ash (seam I) may be characterized as follows*: SiO_2, 56.26 per cent; Al_2O_3, 20.67; Fe_2O_3, 9.45; CaO, 4.08; MgO, 2.42; SO_3, 3.92. Melting temperature, 1290°C.

A preliminary petrographical study has shown that the predominant component of the coals is clarain composed of a cuticule and tar bodies; durain in these coals has a similar structure. Vitrain is encountered in small bands or in lenses with a well-pronounced cellular structure.

A semi-coxing of coals from seam I made in a Fischer retort showed taht the recovery of primary tar varied within the range of 6—8.5 per cent to dry coal; of liquor, 8—10.85 per cent; of gas, 7—10; of semicoke, 71—79; the semicoke is not caked. The primary tar is characterized by a high content of phenols, amounting to 50—60 per cent.

The author has made a study of the liquor and the tar produced in the distillation of coals of seam I in a larger retort of the semi-plant type.

Table 2

Sample	S^r	H^r	N^r	O^r	$\frac{O+N}{H}$
Pack 1, from the shaft	75.67	5.31	1.76	16.89	3.5
Packs 1, 2 » » »	74.8	5.23	2.20	16.33	3.5
Packs 4, 5 » » »	75.55	4.98	2.11	16.38	3.7
Packs 7, 8, 9, 10, 11 from the shaft	72.80	5.38	2.71	18.22	3.9
Semi-lustrous coal	75.72	5.35	1.45	16.47	3.35
Lustrous tar-like coal	75.05	5.16	1.42	17.69	3.7

The liquor gives a weakly alkaline reaction and contains 1.16 g/l of ammonia and 16.7 g/l of phenol-cresols (the analysis was made by the method of Coppeshar-Show).

By fractional distillation of the tar, the recoveries of separate fractions were determined: I fraction up to 170°—11.8 per cent to dry tar. II fraction 170—230°—26.33 per cent to dry tar. III fraction 230—270°—10.31 per cent to dry tar.

In the fractions separated out, the content of phenols and of pyridine bases was analysed.

Thus, the content of phenols in the first three fractions is 29.32 per cent to the tar, and the total recovery of phenol-cresols to the coal amounts to 3—4 per cent.

* The analyses of the ash composition were made by the All-Union Heat and Power Engineering Institute.

Table 3

	Content of phenols		Content of pyridine bases	
	in % to the fraction	in % to the tar	in % to the fraction	in % to the tar
I fraction	44.9	5.3	10.36	1.22
II fraction	74.5	19.6	5.42	1.43
III fraction	42.92	4.42	2.92	0.30

The primary tar is a good foaming agent in the concentration of the Achisay lead-zinc ores, owing to which its use for this purpose is of interest, since the Achisay and Margalimsay concentration mills of the People's Commissariat for Non-Ferrous Metals are situated close by. The Achisay polymetallic combinate is at present organizing the construction (after the author's design) of a small coal-distillation plant (with an output of 40—50 tons of coal per 24 hours); it implies a full utilization of the tar as a flotation reagent for concentration mills (instead of butyl xanthogenate) with a possible separation in the future of light motor fractions. The high phenol content of the tar makes it valuable as a raw material for producing antiseptics: creoline, etc.

The value of the Tas-Komyrsay coals, however, lies not only in the adequateness of their chemical treatment, but in the efficiency of their utilization as fuel too. From the standpoint of quality (moisture content, ash content, calorific value, stability when stored, etc.), the Tas-Komyrsay coals are among the best, if not the very best, of the Jurassic coals of Kazakhstan, and are doubtless superior to the coals from the Lenger deposit *, at present worked with a view of meeting the requirements of South Kazakhstan and of the adjacent regions of Central Asia.

The geological outlook for the future of the Tas-Komyrsay deposit is very favourable. The deposit has been ready for exploitation which is limited only by the absence of railway communication. A realization of the latter will enable the deposit to be made a reliable fuel base of South Kazakstan. The route reconnoitry has been already made.

The high recovery of phenols in the distillation of the Tas-Komyrsay semi-stone coals makes them valuable also as raw material for the chemical industries, particularly in meeting the local requirements.

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* According to the author's estimates 1 kg of the Tas-Komyrsay coal can replace when burnt up to 1.5 kg of the Lenger coal.

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GEOLOGY

THE CASPIAN TERRACES OF THE SOUTH MANGYSHLAK SHORE

By P. V. FEDOROV

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Along the shore of South Mangyshlak stretches a recent Caspian terrace—a beach up to 1.8—2 m high* with *Mytilaster lineatus* Gm. As this mollusc appeared in the Caspian Sea in 1918—21, and the highest sea level since that time was reached in 1929, this shore line may be referred to 1929.

Higher up there is located a complex of coastal forms with *Cardium edule* L. At the steep shores an abrasional terrace is observed, 6 m in height; at the beachy shores there are widely developed: an accumulative terrace of the same height, shore barriers, bars and spits, forming lagoons cut off from the sea, many of which have already dried up.

It is characteristic that nearly all of the bars and spits have a foundation of old Caspian sandstones and shell-rocks lying in the form of an embankment at their base.

The shore line of the terrace with *Cardium edule* throughout the coast does not rise above 6 m. Separated by a gently sloping bench, higher up more ancient terraces are located. The lower, 12 m high, preserved on low shores and in bays, is composed of friable oolitic sandstones with an abundant fauna of large massive *Didacna praetrigonoides* Nal., *D. parallella* Bog., *Dreissensia polymorpha* Pall., *Theodoxus pallasi* Lindh. These deposits compose also the tops of the old Caspian foundation at the base of the bars and are exposed at the bottom of the lagoons. The mode of occurrence, cross-bedding and outlines in the plan permit to regard them as ancient bars and spits.

The upper terrace, 24—25 m high, almost everywhere an abrasional one, is distinctly visible on Tyub-Karagan, widely developed on cape Melovoy and may be traced at the same height further towards Kenderli bay.

In the sands and pebbles covering the terrace there are found small and thin-valved *Didacna praetrigonoides* var. *cristata* Bog., *D. catillus* var. *elongata* Andr., *D. protracta* var. *oblonga* Andr., *Dreissensia polymorpha*, etc. Referring the forms of the *D. praetrigonoides* group to the Khvalynian, the following may be distinguished: an e a r l y K h v a l y n i a n abrasion terrace (25 m), formed at the epoch of transgression, and a l a t e K h v a l y n i a n accumulative terrace (12 m), as well as a series of oolitic sandstones and shell-rocks, lying at the base of bars and spits formed at the time of a regression of the sea**.

* The marks of shore (abrasion) lines are given in meters above the Caspian Sea level (—27.3).

** All the conclusions as to the direction of the movement of the bottom material and attending washing or accumulation of shores, depending on the shifting of the shore line, are based upon the application of the theory of alluvial action of waves (3).

Higher up, being separated from the Khvalynian terraces by a sharp abrasion bench up to 18—20 m in relative height, there is located a series of more ancient terraces at the respective height of 43, 51; 57, 65, 67—70 and 70—73 m. These levels are distinctly seen in the region of the bay of Alexander-Bai and on the shores of the Karaghie (Batyr) depression; a former bay of the ancient Caspian Sea. It is only on terraces on the 67 m and lower levels that occur sparse whole valves of *Didacna* from the *D. delenda* Bog. group, which on the terraces of 67—70, 65 and 57 m have a peculiar appearance. These are small shells with a large number of costae, a sharply prominent anterior and somewhat reduced posterior edge. On the terraces of 51 and 43 m typical *D. delenda* are encountered. Lower down the slope and at the base of the early Khvalynian terrace (by which they are covered) there occurs a series of shell-limestones with pebbles and a large number of large massive *D. delenda*; *D. incrassata* Bog., *Dreissensia rostriformis* Desh.; *Theodoxus pallasi*; *D. surachanica* Nal. non Andr. These deposits stretch in a belt along the shore. On the beachy shores they build up gently sloping embankment at the base of the bars and are sometimes exposed on the bottom of the lagoons (in both cases under late Khvalynian).

Scheme of the Caspian Terraces of the Mangyshlak
(Tyub-Karagan and the coast between cape Melovoy and Kenderli bay)

Name of terraces and morphology		Height above Caspian Sea level, m	Principal fauna and its character	Variations in salinity	Shifting of shore line and development of the coast	Principal events of the Q period	
Recent Caspian [Q_2cd_2]		1.8—2	*Mytilaster lineatus* Gm.	A certain increase in salinity	Regression, increase of bars and spits	1929	Post-glacial epoch
New Caspian [Q_2cd]	Bars [Q_2cd_2]	3—5	*Cardium edule* L.	A certain increase in salinity	Regression, increase of bars and spits		
	Terrace [Q_2cd_1]	6		A certain freshening	Transgression, abrasion	Atlantic time	
Late Khvalynian [Q_1hv_3] (terrace, bars)		12	*Didacna praetrigonoides* Nal. (large, massive)	A certain increase in salinity	Regression, increase of bars and spits	Late-glacial time. Evaporation predominates	
Early Khvalynian [Q_1hv_1] (abrasion terrace)		25	*D. praetrigonoides* var. *cristata* Bog. (small, thin-valved)	A certain freshening	Transgression, abrasion	Flow of thawing waters of the Valday glaciation (Würm) of plains and mountains	
Late Khazarian [Q_1hz_2] (terrace, bars)		20—25	*D. delenda* Bog., *D. incrassata* Bog., *D. surachanica* Nal. (large, massive)	A certain increase in salinity	Regression, increase of bars and spits	Dnieper-Valday (Riss-Würm) interglacial epoch. Evaporation predominates	
Early Khazarian [Q_1hz_1] (abrasion terraces)		70—73 (60—70, 65, 57, 43)	*D.* ex gr. *delenda* Bog. (small, thin-valved)	A certain freshening	Transgression, abrasion	Flow of thawing waters of the Dnieper glaciation (Riss) of plains and mountains	

The outlines in the plan; mode of occurrence; characteristic cross-bedding and lithological composition point to the existence of a system of bars and lagoons more ancient than the late Khvalynian.

Referring the complex of forms of *D. delenda*, *D. incrassata*; *D. surachanica* to the Khazarian; the following may be distinguished: the early Khazarian abrasion terraces at the heights of 70—73, 67—70, 65; 57, 51 and 43 m with characteristic *D.* ex gr. *delenda*, formed at the time of a transgression of the sea, whereas the late Khazarian accumulative forms—series of shell-rocks, systems of bars and spits, containing a rich fauna of large massive *D. delenda*, *D. incrassata*; *D. surachanica*—formed at the time of a regression.

Consequently; there are three systems of bars and lagoons formed at the epoch of regressions: the late Khazarian, the late Khvalynian and the phase of a regression of the sea with *C. edule*, each subsequent system having developed at the place of the preceding one, repeating its forms and outlines.

Above 70—73 m no ancient shore lines and deposits of the Caspian have been discovered. Some forms closely approaching *D. rudis* var. *catillus-rudis* Nal.; found in the Karaghie depression, furnish some grounds to assume here the presence of the Baku deposits covered by the early Khazarian transgression.

An analysis of the morphology and fauna of the terraces permits to distinguish three ancient coast complexes throughout the coast of South Mangyshlak; which not only have morphological boundaries; but also contain a characteristic fauna.

I. The Khazarian Complex. Early Khazarian terraces [Q_1hz_1], mainly abrasional, at the heights of 70—73 to 43 m, whose morphological continuation is a series of shell-rocks, lying below 30 m (down the slope) and forming along the coast a system of bars and lagoons, produced at the period of late Khazarian regression [Q_1hz_2]. It is characteristic that small *D.* ex gr. *delenda* occurring on the terraces of 67—57 m, lower down the terraces are gradually replaced by larger forms, attaining in the late Khazarian deposits of bars the greatest size, thickness of valves and richness of forms (*D. delenda*; *D. incrassata*; *D. surachanica*).

II. The Khvalynian Complex, consisting of an early Khvalynian terrace [Q_1hv_1], 25 m high, the most frequently an abrasional one, and of an accumulative twelve-meter terrace, a system of bars and lagoons of the period of late Khvalynian regression [Q_1hv_2], which are morphologically connected with the former. A regular increase in the size of the valves of *Didacna* is also observed down the terraces.

III. The New Caspian Complex (with *C. edule*) [Q_2cd] consists of a six-meter terrace, marking a phase of transgression, and a system of bars and spits 3—5 m high; formed during a regression of the New Caspian Sea. Here no regular increase in the size of the shells from the upper terraces down to the lower ones is observed; on the six-meter terrace, however, more frequently relatively small *C. edule* and *Didacna* are found, and in the deposits of the bars, larger valves of these.

The heights of all the terraces and their fauna persist throughout the coast.

Thus, here there are traces of three stages of development of the coast, each of these consisting of phases of transgression and regression, with which changes of the fauna complexes are connected.

Great interest belongs to the regular increase in the size and thickness of the valves of *Didacna* from the upper terraces to the lower ones within each complex, and the connexion between the changes of the fauna complexes with the phases of regression, since each transgression would bring a new fauna. This close association of the development of faunistic complexes with the shiftings of the shoreline (variations of the level) justifies, as it seems to be, the assumption of a relation of the increase in the size of shells to the improvement of the ecological conditions of the environment (2)—a certain increase in the salinity and, possibly, warming of the climate, and of the decrease of the shell size—to a certain freshening of the sea and cooling. The transgressions are thus accounted for by an increase of the water mass of the sea at the expense of an increase of the supply of waters from the bands; and the regressions, to a decrease in the supply and a predominance of evaporation.

The author does by no means intend to underestimate the influence of tectonic movements upon the course of the geological history of the Caspian Sea; he only believes following the view of Andrussov ([1]) that, along with these movements and some other factors, the principal part was played by climatic conditions; otherwise it would be impossible to account for the intimate relation between the changes of the faunistic complexes and the transgressions and regressions.

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GENETICS

ONTOGENESIS OF THE SEX DIFFERENCE OF SUSCEPTIBILITY IN *DROSOPHILA MELANOGASTER*

By P. G. SVETLOV

(Communicated by L. A. Orbeli, Member of the Academy, 27.IV.1943)

In our preceding note (1) it has been shown that males of *Drosophila melanogaster* are more susceptible to the influence of injurious factors (including starvation) than the females. In order to inquire into the nature of the sex difference of susceptibility, it is necessary in the first line to settle the question, whether it is characteristic of the imaginal instar only, or may be detected at earlier stages of ontogenesis as well. If the sex difference of susceptibility is a secondary phenomenon, linked e. g. with the different behaviour of males and females, sexual activity, etc., it should be confined to the imaginal stage only. And contrariwise, if this phenomenon is directly underlain by hereditarily determined properties of the cell's protoplasm, then *Drosophila* may be expected to display it at any phase of its life-cycle, since it is well known that in *Drosophila* the sex is determined at the moment of fertilization; and that the chromosome sets in the cells of the two sexes are different at all stages of development.

In order to clear up this question experiments were carried out with larvae and pupae of *Drosophila melanogaster*. The larvae were tested for the effect of starvation, cold and ether vapour. It stands to reason that with pupae experiments involving starvation are impracticable. But as our preceding experiments had shown susceptibility to starvation to be blocked with susceptibility to any other harmful factor, we find it possible to restrict ourselves in this case to the effect of ether and chill. The ether test proved especially convenient.

In order to bring out the sex difference of susceptibility, the two following methods were used: 1) One portion of larvae and pupae from the same hatch were subjected to treatment, the intensity of the latter being so chosen as to ensure the death of a considerable part of the larvae and pupae before they would reach the imago-stage. The other portion from the same hatch served as control. Upon hatching of the flies, the ratio between the males and females was compared in experiment and control. 2) Such individuals as perished at the late pupal stage were taken out of the puparium and their sex was determined. The number of males and females under experiment could thus be established. 3) In larvae the sex was determined prior to experiment by the size of their sex anlages. In spite of its apparent advantages; this method is not exempt of a number of drawbacks; it was therefore applied not without qualification.

L a r v a e. Twelve satisfactory tests were carried out with larvae: seven involving starvation; three, ether; and two, chilling (from -1 to $+5^{\circ}$). When deprived of food the larvae survived for a 24-hour day, or somewhat longer. The ether treatment lasted 12—15 minutes; when saturated; chilling, up to two 24-hour days.

All the tests gave clear-cut negative results. The experiments comprised a total of 1263 larvae, the number of controls being 1298. Upon hatching, six of the experiments gave a small prevalence of females in the controls, while in six others the relations were reverse. The total ratio of hatched males in control and experiment approaches 50 per cent (Table 3). It can thus be considered that the larvae show no sex difference of susceptibility at all.

P u p a e. The pupae were classed into three age groups: 1) young pupae; up to 48-hour-old; 2) pupae with transparent red eyes, and 3) darkened pupae with well perceivable wings. Young pupae were found to be much more susceptible: the dose of ether given to them (when saturated) was from 10—15 min., while pupae of older age group were able to survive in ether vapour up to 20—25 min. Data on young pupae are summarized in Table 1.

Table 1

No.	Control					Experiment					Control	Experiment
	Taken	Hatched total	♂	♀	Perished	Under test	Hatched total	♂	♀	Perished		
P Act 5	40	28	9	17	12	39	10	2	8	29	32.14	20.0
P Act 11	70	66	27	39	4	70	16	5	11	54	40.90	31.25
P Act 14	71	62	35	27	9	64	2	0	2	62	56.45	0
P Act 23	106	92	43	49	14	107	20	9	11	87	46.72	45.00
P Act 24	129	122	51	71	7	130	16	5	11	114	41.80	31.46
P Act 25	187	186	77	109	1	137	35	14	21	102	41.91	40.00
P Act 26	187	186	77	109	1	127	95	50	45	32	41.39	52.62
P Act 28	88	86	49	37	2	64	49	21	28	15	56.97	43.38
Total	691	642	291	351	49	738	243	106	137	495	45.34±1.76	43.62±3.16

At the first glance at this table the result seems to be a negative one: the total percentage of hatched males and females is almost coincident in control and experiment. However, in all experiments except one (P Act 26) the per cent of hatched males was found to be lower than in the control, i. e. there is a regular deviation in favour of a higher susceptibility in males. The experiment was run in two replications with previous sex determination in the larvae. In one of these experiments the percentage of hatched males and females was found to be almost equal (280 and 270), in another there was a sharp prevalence of females (28 and 90 per cent, respectively). This vagueness of the results is likely to have an objective significance: the sex difference of susceptibility does only appear at those stages and is not yet clearly expressed.

The pupae of the second age group (red-eyed) show a clear-cut sex difference of susceptibility. There were carried out five experiments embracing 316 pupae. In all experiments the percentage of hatched males was considerably below that of the females. The total percentage of hatching is given in Table 3.

Pupae of the last development stage (third age group, comprising 525 pupae) were tested in seven runs. All of them gave positive results. In order to make the results more obvious, in four of the experiments all the dead pupae have been taken out of the puparia and their sex determined. Out of the 310 pupae of this group only 15 defied precise sex determination. Neglecting these, it could be established that the 295 flies of this experiment consisted of 135 males and 160 females. The results of these experiments are given in Table 2. The excess of males that perished in these experiments has been found by a direct count of the dead flies, instead of the indirect method previously applied. The figures of Table 2 do not leave space to any doubt as to the presence of a well-pronounced sex difference of susceptibility in pupae of this age.

By comparing the data obtained from different development stages we receive an explicit answer to the question posed at the beginning of this article. The results

Table 2

No.	Under test		Hatched		Perished	
	♂	♀	♂	♀	♂	♀
P Act 6	42	54	26	46	16	8
P Act 9	42	36	5	14	37	22
P Act 12	22	33	15	28	7	5
P Act 32	29	37	6	13	23	24
Total	135	160	52	101	83	59
Per cent	100	100	31.11± 3.98	63.12± 3.69	68.89 ±3.98	36.88 ±3.69

become still more illustrative if the data obtained in both control and experiments be compared to the expected normal ratio of males to females, which under normal conditions should approach 1 (50 per cent of males to 50 per cent of females). These data are given in Table 3.

Table 3

Development	Number of flies		Per cent of flies hatched		Difference between results to be expected and actually observed	
	control	exper.	control	exper.	control	exper.
Larvae	1293	1263	47.7	56.9	2.3±1.14	−0.9±2.8
Pupae 1st stage	691	732	45.34	43.62	4.66±1.76	6.38±3.06
» 2nd »	273	316	51.31	35.57	−1.69±2.90	14.43±3.91
» 3rd »	528	525	46.99	36.03	3.01±2.02	13.97±2.88

It may be seen that in the control the difference between the expected normal and the actually observed is quite insignificant, scarcely exceeding the value of the probable error. In experiments with larvae this difference varies within the same range as in the control. In pupae of the first instar it is somewhat greater, but its significance is statistically disputable. In pupae of the second and third instar this difference assumes much larger values, and its significance is beyond any doubt.

Thus, the sex difference of susceptibility arises during the ontogenesis, and not leapwise with this, and augments after it has arisen. On the one hand, it is not based in males and females of *Drosophila melanogaster* directly on the presence of a male or female set of chromosomes in the cells of their bodies, for it is lacking in larvae. On the other hand, inasmuch as it is not confined to the imaginal instar alone, but may be traced rather far back in the ontogenesis, it should be considered to be one of the important characters of sex differentiation. The susceptibility difference found is obviously referable to primary susceptibility. The latter is known to be intimately linked with the intensity of metabolism, and should be believed to be essentially determined by the properties of the biochemical structure of the proteins making up the body cells.

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MICROBIOLOGY

ON VOLATILE ORGANIC COMPOUNDS EVOLVED BY LIVING ORGANISMS AND ASSIMILATED BY SOIL MICROBES

By N. G. CHOLODNY, Member of the Academy of Sciences of the Ukrainian SSR

1. The atmosphere has long since been known to contain insignificant admixtures of various organic substances in the form of gas or vapour. All these substances seem to be of biogenic origin, and may be classed into three groups: 1) gases and vapours arising from deposits of organic matter: products of life-activity of organisms of the past geological epochs (oil, natural gases); 2) gases and vapours formed of dead remains of recent organisms disintegrated by soil and water micro-organisms; 3) volatile organic substances released into the atmosphere by animals and plants in the process of their life-activity. The third group of substances is the least studied both with respect to quantity and chemical composition, and to the part which they play in Nature, as a possible source of food for other living beings, in the first line, for micro-organisms. This note contains a description of the results of some experiments made along this line; unfortunately, these experiments have remained unachieved becaused of the inroad of the German invadors into the Ukraine.

Research work carried on in the course of 1939—1940 at the microbiological laboratory of the Kiev University brought out the wide occurrence in soil of microbes capable to assimilate from the surrounding atmosphere the most various organic substances, including such as are insoluble in water. In search for the natural sources of this «aerial nutrition» of the soil micro-organisms I have been led to assume that live roots of higher plants, their germinating seeds, tubers, bulbs and other underground organs, as well as the mycelium of fungi, larvae of insects and other soil inhabiting animal and vegetable organisms may possibly evolve in the process of their metabolism a certain amount of volatile organic compounds. Along with the gaseous products of various kind of fermentation, these volatile substances may be assimilated by the soil microbes directly from the air which is contained in the soil. In order to check up this assumption, use has been made of the previously described(1) method of growing soil dust on the surface of a clean glass.

In small glass cylinders supplied with ground-stoppers a few cover-glasses were suspended on nickel wires; one side of these glasses was covered with a layer of soil dust. Somewhat below the glass a gauze net was spread containing germinating seeds, while some water was poured over the bottom of the cylinder.

The cylinders were usually exposed to scattered light in order to allow the overground parts of the sprouts to assimilate carbon dioxide and thus fill up the provision of oxygen in the receiver. Moreover, ventilation was practised from time to time, by removng the stoppers for a short while (several seconds). The distance between the glasses with dust and the seeds was so chosen as to prevent any contact between the developing sprouts and the glasses.

Control preparations of soil dust taken from the same soil sample were put into identical cylinders; less the seeds. The experiments were carried on during summer

season, at an indoor temperature of about 25° C. After some four—five days the preparations were taken out of the cylinders, dried, fixed and stained with gentian-violet. Sometimes they were also examined *in vivo*.

2. By comparing the controls with the experimental preparations it could easily be made evident that in those cylinders which contained sprouts the soil micro-organisms usually developed much better than without sprouts. In some cases the difference was a quite conspicuous one: when estimated visually, the total increment of the «live mass» of micro-organisms in the experimental preparations might well be tenfold and even hundredfold that of the controls. While in the latter bacteria, actinomycetes and fungi formed but scanty accumulations in the closest neighbourhood of the soil clots which were their unique source of food, the experimental preparations offered an opulent growth of micro-organisms in the intervals between the soil particles, too. The whole area of the glass used to be covered almost continuously with gigantic colonies of well fed cells. Strikingly enough, there was always one form prevailing, occasionally, two or three forms.

The results of the experiments with germinating peas were especially illustrative. The other *Leguminosae* also brought about a considerable increase in the growth of microbes. This should possibly be explained by the evolution of some nitrogenous volatile organic compounds that might have simultaneously served for the microbes as a source of carbon. In experiments involving cereals (oat; barley, wheat) the growth of microbes was less opulent; nevertheless, preparations from experimental cylinders invariably showed a considerably higher increment than the controls.

Along with these experiments, others were made using larger objects; such as potato-tubers, and also apples. Now glass bells served as receivers, put upon plates filled with water. Ripe apples should be mentioned here, although they have no direct bearing on the object treated here; but in their presence the growth of actinomycetes went on very intensely. However; a similar result was obtained in some experiments with cereals, too. *Leguminosae*, especially peas; promoted the development of large, rod-shaped bacteria which formed long threads on the preparations; occasionally uniting into thick winding strands.

3. All these experiments with their clear-cut results put it beyond any doubt that many of the soil micro-organisms are liable to assimilate some volatile organic compounds evolved by germinating seeds and some other vegetable organs.

The assumption that the substances responsible for the abundant growth of microbes in the tests here described had arisen in the process of bacterial disintegration of the complex organic compounds contained in rotting seeds; should be rejected: the sprouts used in our experiments were thoroughly selected, remained healthy throughout the test, and, owing to the precautions taken, their development within the glass cylinders had an entirely normal course.

The plants seemed to experience no oxygen deficiency. Yet, in the case of sprouts with thick cotyledons, viz. in peas, the possibility of intramolecular respiration cannot be excluded; but this can take place also in soil, under natural conditions of germination.

Experiments in which germinating seeds were replaced by green leaves severed from the plant, yielded negative results. After all, this was not surprising, for it seemed rather improbable that newly synthesized organic substances be released at once.

We are thus left to suppose that in the experiments described the main course of the formation of volatile organic compounds consisted in the transformations of the nutrient substances stored by the seed and, possibly, some metabolic processes taking place in the roots of the plants. The latter seems especially probable for cereals whose sprouts are endowed with a well developed root system; while their endosperm is rather soon evacuated.

4. The evolution of volatile organic compounds; including nitrogenous ones, by living underground plant organs (germinating seeds, tubers, roots), as revealed by these experiments; has to my knowledge never been reported in literature.

Yet; the quantity of these compounds; if we were to judge by the intense growth of the soil micro-organisms feeding on them; should in some cases attain appreciable figures; and if we take into account the huge mass of seeds; tubers and other plant organs germinating in the soil, as well as that of the live roots it is pierced with; it may safely be maintained that this food store must necessarily play an important rôle in the life of the soil microflore; and that it should not be neglected when forming a general quantitative estimate of the circular exchange of matter in the biosphere.

From the experiment on apples it seems probable that those volatile organic substances which; being evolved in insignificant amounts by flowers, fruits and other overground plant organs; are sometimes responsible for their odour, do not always dissipate without leaving any traces in the surrounding atmosphere. The conditions being favourable; they may also be absorbed and assimilated by other organisms. The ripe fruits covering the ground under trees and shrubs pay their tribute not only in the form of those substances that are washed out of them by water, but also which are diffused around in the form of easily penetrating molecules of gases and vapours.

Indeed, it does in no way contradict the assumption that the principal biological rôle of the volatile substances evolved by flowers, fruits and leaves consists in the oligodynamic effect produced by them upon other organisms, as a means of defense, attraction, etc.

Ripe apples are known to evolve ethylene. It is possible that this gas also serves as a source of carbon for some soil microbes.

Conclusions

1. Germinating seeds of various plants release into the surrounding atmosphere some volatile organic substances. In some cases (for example, in the case of peas) these substances seem to contain nitrogen.

2. It seems highly probable that the roots in sprouts advanced in their development also take part in the emitting of volatile organic compounds.

3. The volatile organic substances evolved by sprouts are absorbed and assimilated by the various soil micro-organisms, chiefly by bacteria and actinomycetes.

4. The volatile organic compounds evolved by fruits, for example, by ripe apples, may also serve as a source of food for some micro-organisms.

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MICROBIOLOGY

TERATOLOGY OF THE OVULE IN *COMPOSITAE*

By V. L. RISCHKOV

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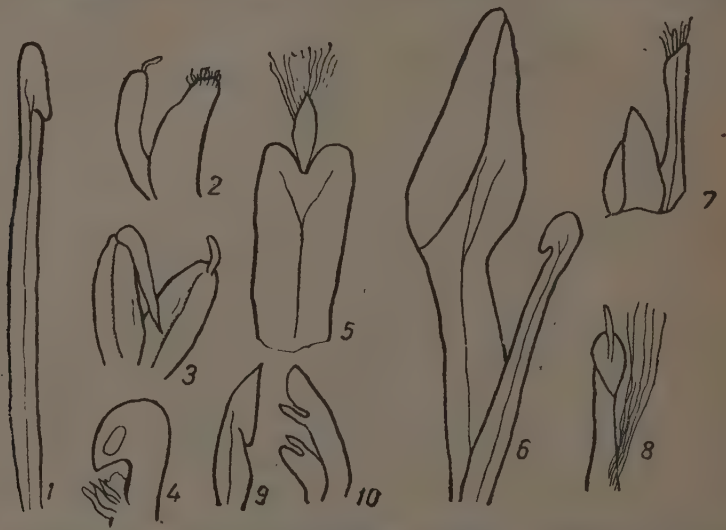
In a series of papers I have endeavoured to reveal the regularities underlying teratological phenomena. The vast material furnished by cases of mass virescence of flowers which are largely due to filterable viruses, has made it possible, I believe, on the one hand, strictly to prove the fitness of teratological material for the solution of problems of comparative morphology, and, on the other, to show the significance of purely ontogenetic regularities in the development of teratological phenomena and to disclose their rôle here as a factor that obscures the phylogenetic picture.

One of the difficult and important problems in comparative morphology is that of the origin of the ovule. I was able to show that the behaviour of ovules in the case of antholysis is utterly different in different families. The families having a marginal placenta, such as *Cruciferae*, for instance, display on antholysis an obvious connexion between the ovules and carpels. Even if the ovary has completely disintegrated into two carpels, along the edges of these carpels ovules or their derivatives may sometimes be found. In families with an axial placenta, such as belong in the order *Personatae*, the ovules are not usually connected with the carpels in the case of antholysis, but develop instead on stem-like organs. Their behaviour then suggests an axial origin. It is only because I had at my disposal a very large accumulation of data on the flower modifications caused by stawboor, a virus disease, that I was able to find in *Atropa belladonna* such antholysed flowers in which the connexion between the stem-like placentae and the carpels was unmistakable. From this I felt entitled to conclude that in the representatives of *Personatae* the ovule is a derivate of the carpel. It is not possible to ignore, however, the different behaviour of ovules in the families with a marginal and with an axial placenta, which I have regularly observed on an abundant material. This phenomenon cannot possibly have been a mere coincidence, since it has been repeatedly described by various authors in various families, and not a single exception has ever been reported. I tried to account for the existence of these two types of antholysis by a different origin of the placenta, taking it for granted that the marginal one has arisen from the marginal situation of sporangia, and the axial as a results of the branching of the carpel similar to that which occurs with sporiferous leaves of forms in the fam. *Ophioglassaceae*.

Among a great many data on antholysis in *Compositae* I have found nothing that might in any way contribute to the understanding of the origin of the ovule. As a rule, a plant of this family, when affected by antholysis, either has its ovule transformed into a leaflet which grows, ruptures the ovary and penetrates outside, or it shows a proliferation of the flower, while the ovule disappears altogether.

The morphological series of gradual transitions bridging from the transformations of a normal organ to extreme deviations from the norm, and constituting the main value of teratological findings, appear to be a great rarity in the sphere of the female generative apparatus of *Compositae*. None the less important for that are the diverse transformations of the ovules I have discovered in studying the kok-saghyz yellow, both in plant and in other genera and species of *Compositae*.

But it is the kok-saghyz that affords the richest material for the study of the various transformations of the ovule (figure, *5* and *8*). These transformations end in the ovule, being replaced by a long leaflet which breaks ovary and gets outside. In other heads we meet with flowers having their ovule on a more or less elongated pedicle. Usually these ovules, however long the pedicle that supports them, retain traces of the embryo-sack; as the ovule is further reduced, it only forms a bend in



Transformation of ovule in the yellow affected plants of kok-saghyz and other *Compositae*: *1*) ovule of *Taraxacum* on a long pedicle; *2—3*. ovule pairs from ovary of *Cichory*; *4*, ovule of *Matricaria inodora* with growing cone at its base; *5—8*, ovules of kok-saghyz described in this paper; *9—10*, ovule and a pair of ovules from ovary of *Chrysanthemum* flowers

the upper part of the pedicle and the whole looks like a hook. It may be remarked that the elongated pedicle supporting the ovule forms a groove and has a dorse-ventral symmetry, that is, assumes a more or less leaf-like appearance. Further transformations of the ovule towards the shape of a leaflet may then develop in different forms. Very often malformed leaves are found inside the ovary, which in their upper portion bear a thin and short cylindrical process and, not infrequently, also trichomes. This process either sets in the depression formed by the leaf or branches from the midrib. One gets the impression as if the leaf-like form is assumed by the pedicle when the ovule itself is reduced to the process. In cases when the ovule is replaced by a large leaflet, one sometimes finds on it this process branching from the midrib.

When there is a proliferation of the flower and the growing cone appears inside the ovary, the ovule or its derivate, the leaflet with the process may be retained still (figure, *7*). In such cases the ovule looks as if it were the first leaf of the growing cone. Several hundreds of such flowers have been examined by me. In other cases the picture is complicated by the development of a peculiar leaflet alongside of the ovule. This leaflet is turned to the ovule with its morphologically upper side, and to the growing cone, with its under side (figure, *6*). Thus, the ovule and the leaflet are not so facing each other as the adjoining leaves of a single bud do, but rather

have their upper sides in opposition. I look upon this leaflet as a second ovule which does not develop in *Compositae* under normal conditions. In favour of this view speaks the likeness of the innervation in the ovule and in the leaflet; and what is a more convincing argument, I have observed the appearance of two genuine ovules on the described material in other species of *Compositae*. On this observation I will dwell below.

The fact that the continuation of the flower axis may at once be observed in the growing cone and in the derivates of the ovule, speaks against the axial origin of the latter organ. The simultaneous occurrence of the growing cone and the ovule in the ovary of affected flowers has also been stated in *Matricaria inodora* (figure, *4*). *Cichorium intibus* at a certain stage of antholysis also retains ovules in its lower ovary. Here the reduction of the ovule can be followed up through every kind of transition. The pedicle increases strongly in length and thickness to a sausage-like organ which terminates in a reduced ovule in the form of a beak turned downwards. This series is closed by the sausage-like pedicles of the ovule with a small withered process. It is important to point out that very often flowers with two ovules are encountered (figure, *2* and *3*). In some cases both of them are equally well developed, in other cases one is more strongly reduced that the other. When affected by antholysis, as described here, the species of *Chrysanthemum*, too, have their ovule thickened with a certain reduction of its generative part, and there are many ovaries with two ovules in each (figure, *9* and *10*).

This regular appearance of the second ovule in *Compositae* suffering antholysis is a very remarkable fact, indeed, for it brings to light the development of phylogenetically earlier phases in teratological formations. If the transformation of a flower into a vegetative shoot may as yet be considered as a retrograde development, or such as gives rise to organs normally belonging to the plant, the appearance of a second ovule in a plant whose pistil is formed by two carpels, but only one ovule has been retained in the process of phylogeny, is certainly a case of true atavism in the sense in which this word was meant by Darwin. Herein lies primarily the theoretical value of the finding. It consists further in revealing once more the regular character of teratological phenomena. He who denies their significance in phylogenetic comparative morphology is at a loss to explain why the pistil in the case of antholysis should disintegrate, as it always does, into a number of leaves equal to the number of carpels, or the appearance of a second ovule due to antholysis be observed in widely different species of *Compositae*.

Among the described material great interest belongs also to the appearance of trichomes in the growing cone inside the ovary of a kok-saghyz flower. These trichomes now arise on the leaves formed from the ovule, and it seems that some portions of such a leaf are reduced to trichomes, now they morphologically replace the leaf as in figure, *8*, where they are seen in front of the ovule transformed into a leaf-like form, which corresponds to the possible development of a leaflet in front of the ovule (figure, *6*). In *Compositae* affected with antholysis the pappus and its derivates turn into leaflets, yet we see that the leaflet too can be replaced by the elements of the pappus, viz. by trichomes. This kind of transformation is apparently a feature peculiar to *Compositae* and testifies against the sceptical attitude of many morphologists toward the probability of homology of elements of the pappus.

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PLANT PHYSIOLOGY

STIMULATION OF POTATO CROPS IN NORTHERN KAZAKHSTAN

By A. A. RICHTER, Member of the Academy, and T. A. KRASNOSELSKAYA

In the summer season of 1942 we were able to try the growing of potato on specially drained gravelly soils of North Kazakhstan by the method of tuber top planting according to T. D. Lyssenko, the tops being cut off from the tubers and stored beforehand. This we compared with the planting of whole tubers. At the same time we extended our experiment to the activation of the tops by chemical stimulants and as such used trace elements which are the most simple and cheap type of salt activator.

The results of these tests may have a practical interest and we feel therefore justified to report them in this paper.

At Borovoe on the grounds of the farm belonging to the health resort administration, a district was alloted for the experiment, cut up into 9 plots. On May 26 1942 these plots were planted with commercial potato varieties (a mixture of White and Early Rose). In one of the plots uncut tubers were buried, while the other received tops stored up for some time before. Of these eight plots one was planted with dry, untreated tops, and the rest with tops that have been soaked for 3 ours in pure water or in a solution of one of the following three chemicals: boric acid, potassium permanganate or zinc sulphate. After soaking the tops were covered with sacking and planted moist the next day.

The quantitative proportion was so chosen that each plot received 110 seed units of nearly equal weight (about 600 g). On all plots the preparation of the seed-bed and treatment during growth were carried out in the usual way (weeding, soil loosening, hilling). The potatoes were harvested on October 1.

In estimating the yield 20 potato plants were chosen on every plot as near in appearance to the average as possible. The weight of their tubers was then recalculated for the respective plot as a whole. In that way the following figures were obtained.

I. Whole tubers (107 hills)	128.4	kg.	Per	100	hills—	120 kg
II. Dry tops (95 hills)	136.8	»	»	»	»	144 »
III. Tops soaked in water (85 hills)	114.8	»	»	»	»	135 »
IV. Tops treated with 0.1% boric acid (98 hills)	125.4	»	»	»	»	128 »
V. Tops treated with 0.01% boric acid (95 hills)	116.9	»	»	»	»	123 »
VI. Tops treated with 0.1% potassium permanganate (82 hills)	121.3	»	»	»	»	148 »
VII. Tops treated with 0.01% potassium permanganate (96 hills)	124.8	»	»	»	»	130 »
VIII. Tops treated with 0.1% zinc sulphate (98 hills)	129.3	»	»	»	»	132 »
IX. Tops treated with 0.01% zinc sulphate (91 hills)	156.5	»	»	»	»	172 »

If these data are expressed in per cent with respect to whole tubers (I) or to dry tops (II), the following figures are obtained:

	I	II
I. Whole tubers	100	83
II. Dry tops	120.5	100
III. Tops soaked in water	113.3	94
IV. » treated with 0.1% boric acid	107.2	89
V. » » » 0.01% » »	103.7	86
VI. » » » 0.1% potassium permanganate	124.1	103
VII. » » » 0.01% » »	108.4	90
VIII. » » » 0.1% zinc sulphate	110.8	92
IX. » » » 0.01% » »	143.3	119

From these figures one may conclude as follows:

First, it is obvious that whole tubers develop a higher ratio of successful hills than top sets do. The viability of the latter is even reduced by their pre-sowing treatment: whole tubers, 97.3 per cent of developed hills; dry tops, 86.4 per cent; soaked tops, 77.3 per cent.

Secondly, the application of a stimulator in the form of a trace element solution goes in some cases to increase the yield of tubers even if the number of developed tops is smaller: with a 0.01 per cent solution of zinc sulphate the ratio of developed hills was reduced to 82.7 per cent, but the yield increased by 19 per cent as compared to that obtained from dry tops, by 43 per cent compared to that from whole tubers. The yield per hectar was 440 centners in the case of whole tubers; 528 centners in the case of dry tops, whereas the zinc-stimulated tops gave 631 centners.

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EXPERIMENTAL MORPHOLOGY

RESTITUTION OF REGENERATION CAPACITY OF A TERRITORY OF THE FORE EXTREMITY BY GRAFTING ON TO ITS WOUND SURFACE PARTS FROM OTHER TERRITORIES

By M. I. EFIMOV

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After an extremity has been severed together with the shoulder girdle, the wound surface becomes covered with skin, and there is no development of parts analogous to those removed. If the skeletal or muscular tissues of the shoulder girdle are retransplanted on to the wound surface, the regeneration territory re-assumes its capacity to form anew both the extremity and the shoulder girdle. The part of the skeleton or muscles of the shoulder girdle may successfully be played by the corium of the territory of the fore limb (1).

The facts established invite our attention to the rôle of the graft and of the tissues it is surrounded with in the restitution of the regeneration capacity of the territory of the fore limb.

Two alternatives have been suggested in respect of this assumption:

1. The graft constitutes the basic part of the regeneration territory which is responsible for the issue of the regeneration process.

2. Being liable to partial decay, the graft call forth the decay of the surrounding tissues, promoting them hereby to the regeneration process. The degree to which the graft is involved into this process may vary largely.

The first assumption receives some support on the part of the work by Umansky (4). From the experimental data Umansky was brought to conclude that any part of the extremity (corium, skeleton, muscles) is apt in the process of its regeneration to reproduce from mesoderm derivatives all the remaining parts.

The second assumption may be reinforced by data of Nassonov (2, 3) who has shown that the regeneration process of an organ can be brought about by grafting small pieces of cartilage under the skin. In the case reported the active rôle in the regeneration process belonged to the local tissues, since the regenerate's structure corresponded to the place of transplantation. The fact that local tissues surrounding the graft partake in the regeneration process is also evidenced by the results of my previous work (1).

In order to check up the assumptions exposed above, special experiments were carried out. The experiments involved replacement of parts of the regeneration territory of the fore extremity, both of the graft and of the surrounding tissues, by parts taken from other territories. As material use was made of white and black axolotls, from seven to twelve months old. The axolotls were kept at room temperature in aquaria, several individuals in each, and fed on meat.

R u n 1. In this experimental run I made an attempt to restitute the regeneration capacity of the territory of the fore limb by way of grafting on to the wound

surface skeletal tissue from the cranium or of the branchial apparatus, i. e. of the skeleton of another territory.

The experimental axolotl was amputated one of its fore limbs together with the shoulder girdle. After several days under the skin of the wound surface was grafted the skeletal tissue of the branchial apparatus (5—6 areas 5—7 mm across) or the skeletal tissue of the skull (area 5—6 mm) from an axolotl of a different pigmentation. After some 20—25 days the area of the skin covering the graft was removed, thus creating a wound surface of a diameter wider than that of the humerus.

Eight axolotls were operated on the whole. In one of them a normal limb has developed. In five the development of an atypical extremity took place; two failed to restitute the extremity. In all cases the regeneration had a skin whose pigmentation corresponded to the place of transplantation. Both the structure of the regenerate and the character of the regenerate's skin testify to the fact that the tissues which surround the graft played the chief part in the course of the regeneration process. As to the part played by the skeleton transplanted, it is possibly confined merely to stimulating the surrounding tissues to the regeneration process.

R u n 2. The only distinction between this run and the preceding one consisted in that in this case I used as a graft the corium of the tail, i. e. from another territory.

The removal of the limb together with the shoulder girdle was occasionally attended by the preservation of the skin of the proximal part of the humerus in the shape of a sleeve. The opening of the skin sleeve was sutured in the shape of a pouch. A few days after rectangular skin area was excised from another axolotl's tail. The epithelial layer was removed mechanically. From the corium of the tail's skin a small many-layered roll was made and grafted into the skin-sleeve of the limb or under the skin of the territory of the fore extremity. A total of fourteen axolotls were operated upon.

In two of the axolotls a normal extremity was found to regenerate. In five the regenerated limb showed atypical features. In eight there was no regeneration at all. In one axolotl the regeneration process assumed a peculiar course, having resulted in the development of a chimeral regenerate. The regenerate attained 8—9 mm, it was cylindrical in shape, with a small bend in the middle; the toe-anlage was missing. It was not until more advanced stage of the process that the regenerate acquired more distinct characters of an extremity. The distal part remained atypical. Its end consisted of a fin-tissue of the tail, inside of which there was a toe-anlage. The nature of this regenerate's structure made it evident that in the regeneration process the leading rôle was held by the tissues that surrounded the graft and were stimulated by the latter to the regeneration process. In one case the graft participated more actively in the development of the regenerate, which is evidenced by the fin-like tissue that formed on it.

The data of both the first and the second experimental runs go to confirm the latter of the two assumptions made to account for the rôle of the graft and of the surrounding tissues in the restitution of the regeneration capacity on the wound area of a fore limb.

Now what is then the part of the regeneration territory which is brought by the graft into an active state? Among the parts by which the graft is surrounded, are the skin of the territory of the fore limb and the trunk muscles underlying the shoulder girdle. There is every ground to believe that the stimulating tissue is that of the trunk musculature.

I have shown in a previous paper (1) that the failure of an extremity to regenerate after it has been severed together with the shoulder girdle is due to the inactive state of the muscles of the wound area. But the skin has a free edge which is in contact with the wound surface, thus enjoying all necessary conditions to partake in the regeneration process.

R u n 3. The aim of this run was to decide about the part of the skin of the territory of the fore limb in the restitution process of the extremity. In order to do so I replaced the tkin of the territory of the extremity by that of the tail. The ex-

tremity was amputated together with the shoulder girdle. After several days the skin of the territory of the limb was replaced by that taken from the tail of an axolotl of a different pigmentation. Upon healing of the skin, I grafted under it the skeleton of the shoulder girdle so as to make the socket of the articulation coincide with the place of the outlet of the nerves. After 20—25 days the skin in front of the articulation socket of the skeleton transplanted was removed; in this way a wound surface with a diameter wider than that of the humerus was created.

A total of fifteen axolotls were operated upon. In four of them small chimeras were formed on the wound surface, i. e. regenerates containing parts of the limb that have developed into a fin-like tail tissue. The pigmentation of the regenerate's skin was intermediate between those of the skin of the graft and of the host. In the other cases of regeneration process there was no limb formed. The results obtained should be interpreted to mean that the skin of the limb territory around the graft is partly responsible both for the arising and issue of the process of regeneration of the extremity.

From the experimental data of the present work the following conclusions may be drawn:

1. Upon a wound surface formed after the amputation of a limb together with the shoulder girdle the deficient parts can be made to develop by grafting on to the wound area the skeletal tissue or the corium of skin taken from other territories.

2. The rôle of the graft is evidently reduced in this case to stimulating the trunk muscles of the wound surface to the regeneration process; for the regenerate's structure is found to correspond in the main to that of the spot of transplantation.

3. The skin of the territory of the fore limb, surrounding the graft, is also responsible for the origination and result of the regeneration process; since in the case it is replaced by skin from the tail, the regeneration process of the organ fails to set in, or also a small atypical outgrowth is formed.

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